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MATHÉMATIQUES

ESTIMATION D'UNE SOMME ARITHMÉTIQUE

Par CONSTANTIN MARDJANICHVILI

(Présenté par I. M. Vinogradow, de l'Académie, le 7.I.1939)

Dans la considération de certaines questions de la théorie analytique des nombres il est nécessaire d'estimer la valeur de la somme

$$T_k^{(l)}(n) = \sum_{1 \leqslant m \leqslant n} \tau_k^l(m),$$

où $\tau_k(m)$ est le nombre des solutions $(x_1, x_2, \ldots, x_k)$ en nombres entiers positifs de l'équation

$$x_1 x_2 \ldots x_k = m.$$

Dans le présent article nous démontrons le théorème suivant:

Quels que soient les entiers positifs $l \geqslant 1$, $n \geqslant 1$, $k \geqslant 2$, *on a l'inégalité*

$$n^{-1} \sum_{1 \leqslant m \leqslant n} \tau_k^l(m) < A_k^{(l)} (\lg n + k^l - 1)^{k^l - 1}, \tag{1}$$

où

$$A_k^{(l)} = \frac{k^l}{(k!)^{\frac{k^l - 1}{k-1}}}. \tag{2}$$

Démonstration. 1. Soit $K(n_1, n_2, \ldots, n_l)$ le plus petit multiple commun des nombres $n_1, n_2, \ldots, n_l$; alors, en observant que

$$\tau_k(m) = \sum_{d/m} \tau_{k-1}(d), \tag{3}$$

nous aurons

$$T_k^{(l)}(n) = \sum_{1 \leqslant m \leqslant n} \Big(\sum_{d/m} \tau_{k-1}(d) \Big)^l =$$

$$= \sum_{1 \leqslant m_l \leqslant n} \tau_{k-1}(m_1) \ldots \tau_{k-1}(m_l) \left[\frac{n}{K(m_1; m_2; \ldots; m_l)} \right] <$$

$$< \sum_{1 \leqslant m_1 \leqslant n} \tau_{k-1}(m_1) \sum_{d/m_1} \sum_{1 \leqslant m_2 \leqslant \left[\frac{n}{d}\right]} \tau_{k-1}(m_2 d) \sum_{\substack{m_q \leqslant n \\ (q=3,\ldots,l)}} \tau_{k-1}(m_3) \ldots$$

$$\ldots \tau_{k-1}(m_l) \left[\frac{n}{K(m_1 m_2; m_3; \ldots; m_l)} \right],$$

d'où en vertu de (3) et de ce que

$$\tau_i(n_1n_2) \leqslant \tau_i(n_1)\,\tau_i(n_2) \tag{4}$$

il suit

$$T_k^{(l)}(n) \leqslant \sum_{1\leqslant m_1\leqslant n} \tau_{k-1}(m_1)\,\tau_k(m_1) \sum_{1\leqslant m_2\leqslant n} \tau_{k-1}(m_2) \sum_{\substack{1\leqslant m_q\leqslant n\\(q=3,\ldots,l)}} \tau_{k-1}(m_3)\ldots$$
$$\ldots\,\tau_{k-1}(m_l)\left[\frac{n}{K(m_1m_2; m_3; \ldots; m_l)}\right].$$

Ensuite nous obtenons de même

$$T_k^{(l)}(n) <$$
$$< \sum_{1\leqslant m_1\leqslant n} \tau_{k-1}(m_1)\,\tau_k(m_1) \sum_{1\leqslant m_2\leqslant n} \tau_{k-1}(m_2) \sum_{d\mid m_1m_2} \sum_{1\leqslant m_3\leqslant\left[\frac{n}{d}\right]} \tau_{k-1}(dm_3) \sum_{\substack{1\leqslant m_r\leqslant n\\(r=4,\ldots,l)}} \tau_{k-1}(m_4)\cdot$$
$$\cdot\,\tau_{k-1}(m_5)\ldots\tau_{k-1}(m_l)\left[\frac{n}{K(m_1m_2m_3; m_4; \ldots; m_l)}\right],$$

d'où, en tenant compte de (3) et (4), nous trouvons

$$T_k^{(l)}(n) < \sum_{1\leqslant m_1\leqslant n} \tau_{k-1}(m_1)\,\tau_k^2(m_1) \sum_{1\leqslant m_2\leqslant n} \tau_{k-1}(m_2)\,\tau_k(m_2) \sum_{\substack{1\leqslant m_q\leqslant n\\(q=3,\ldots,l)}} \tau_{k-1}(m_3)\ldots$$
$$\ldots\,\tau_{k-1}(m_l)\left[\frac{n}{K(m_1m_2m_3; m_4; \ldots; m_l)}\right].$$

En continuant le même procédé, nous aurons enfin

$$\sum_{1\leqslant m\leqslant n} \tau_k^l(m) < n \sum_{1\leqslant m_1\leqslant n} \frac{\tau_{k-1}(m_1)\,\tau_k^{l-1}(m_1)}{m_1} \sum_{1\leqslant m_2\leqslant n} \frac{\tau_{k-1}(m_2)\,\tau_k^{l-2}(m_2)}{m_2}\ldots$$
$$\ldots \sum_{1\leqslant m_l\leqslant n} \frac{\tau_{k-1}(m_l)}{m_l}. \tag{5}$$

2. Nous évaluerons maintenant la somme $T_k^{(l)}(n)$. Il est clair que

$$\sum_{1\leqslant m\leqslant n} \tau_2(m) = \sum_{1\leqslant m\leqslant n} \sum_{d\mid m} 1 = \sum_{1\leqslant d\leqslant n} \left[\frac{n}{d}\right] < n \sum_{1\leqslant d\leqslant n} \frac{1}{d} < n(\lg n + 1), \tag{6}$$

$$\sum_{1\leqslant m\leqslant n} \tau_3(m) = \sum_{1\leqslant m\leqslant n} \sum_{d\mid m} \tau_2(m) = \sum_{1\leqslant d\leqslant n} \tau_2(d)\left[\frac{n}{d}\right] < n \sum_{1\leqslant d\leqslant n} \frac{\tau_2(d)}{d}. \tag{7}$$

En effectuant la sommation partielle, nous trouverons

$$\sum_{1\leqslant d\leqslant n} \frac{\tau_2(d)}{d} < \frac{1}{2} + \frac{T_2^{(1)}(2)}{2^2} + \frac{T_2^{(1)}(3)}{3^2} + \ldots + \frac{T_2^{(1)}(n-1)}{(n-1)^2} + \frac{T_2^{(1)}(n)}{n},$$

d'où, en observant que pour $s \geqslant r$

$$\sum_{2\leqslant m\leqslant n} \frac{(\lg m + s)^r}{m} < \int_1^n \frac{(\lg t + s)^r}{t}\,dt = \frac{(\lg n + s)^{r+1}}{r+1} - \frac{s^{r+1}}{r+1},$$

et en vertu de (6) et (7) il suit

$$\sum_{1\leqslant m\leqslant n} \tau_3(m) < n\left(\frac{(\lg n+1)^2}{2} + (\lg n + 1)\right) < n\,\frac{(\lg n + 2)^2}{2}.$$

En continuant les mêmes raisonnements, nous aurons

$$\sum_{1\leqslant m\leqslant n} \tau_k(m) < n\,\frac{(\lg n + k - 1)^{k-1}}{(k-1)!}, \tag{8}$$

ce qui démontre le théorème dans le cas $l=1$.

3. Supposons maintenant que le théorème est vrai pour $l < l_1$; alors

$$\sum_{1\leqslant m\leqslant n} \tau_k^l(m)\,\tau_2(m) = \sum_{1\leqslant m\leqslant n} \tau_k^l(m) \sum_{d/m} 1 = \sum_{1\leqslant d\leqslant n} \sum_{1\leqslant m\leqslant\left[\frac{n}{d}\right]} \tau_k^l(md) <$$

$$< \sum_{1\leqslant d\leqslant n} \tau_k^l(d) \sum_{m\leqslant\left[\frac{n}{d}\right]} \tau_k^l(m) < n\,A_k^{(l)}(\lg n + k^l - 1)^{k^l-1} \sum_{1\leqslant d\leqslant n} \frac{\tau_k^l(d)}{d};$$

$$\sum_{1\leqslant d\leqslant n} \frac{\tau_k^l(d)}{d} < \frac{A_k^{(l)}(\lg n + k^l - 1)^{k^l}}{k^l} + A_k^{(l)}(\lg n + k^l - 1)^{k^l-1} +$$

$$+ \frac{1}{2} - A_k^{(l)}\,\frac{(k^l-1)^{k^l}}{k^l} < \frac{A_k^{(l)}(\lg n + k^l)^{k^l}}{k^l};$$

$$\sum_{1\leqslant m\leqslant n} \tau_k^l(m)\,\tau_2(m) < n\,\frac{(A_k^{(l)})^2}{k^l}(\lg n + 2k^l - 1)^{2k^l-1}.$$

Par induction nous trouvons ensuite que

$$\sum_{1\leqslant m\leqslant n} \tau_k^l(m)\,\tau_p(m) < n\,\frac{(A_k^{(l)})^p}{(p-1)!\,k^{(p-1)l}}(\lg n + pk^l - 1)^{pk^l-1}. \tag{9}$$

En posant $p=k-1$ et en effectuant la sommation partielle, nous aurons

$$\sum_{1\leqslant m\leqslant n} \frac{\tau_k^l(m)\,\tau_{k-1}(m)}{m} < \frac{(A_k^{(l)})^{k-1}}{(k-1)!\,k^{(k-1)l}}(\lg n + (k-1)k^l)^{(k-1)k^l},$$

d'où il suit que

$$\prod_{1\leqslant l\leqslant l_1-1} \sum_{1\leqslant m_l\leqslant n} \frac{\tau_k^l(m_l)\,\tau_{k-1}(m_l)}{m_l} < \frac{(A_k^{(1)}A_k^{(2)}\ldots A_k^{(l_1-1)})^{k-1}}{((k-1)!)^{l_1-1}k^{\frac{(k-1)l_1(l_1-1)}{2}}}(\lg n + k^{l_1} - 1)^{k^{l_1}-k}.$$

À force des inégalités (5) et (8) nous obtiendrons finalement

$$\sum_{1\leqslant m\leqslant n} \tau_k^{l_1}(m) < n\,\frac{(A_k^{(1)}\ldots A_k^{(l_1-1)})^{k-1}}{((k-1)!)^{l_1}k^{\frac{(k-1)l_1(l_1-1)}{2}}}(\lg n + k^{l_1} - 1)^{k^{l_1}-1},$$

ce qui démontre le théorème.

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MATHEMATICS

ON ONE PROPERTY OF REGULAR SPACES

By D. MILMAN

(Communicated by S. N. Bernstein, Member of the Academy, 17. I. 1939)

I. We introduce the following definitions:

Definition 1. A topological space is called ϑ_0-compact, where ϑ_0 is a limiting ordinal number, if for every non-increasing sequence

$$L_1 \supseteq L_2 \supseteq \ldots \supseteq L_\vartheta \supseteq \ldots \qquad (\vartheta < \vartheta_1;\ \vartheta_1 \leqslant \vartheta_0)$$

of non-empty closed sets there exists a non-empty intersection.

Remark. It is easily seen that 1) an ω-compactness is an ordinary compactness, 2) a compactness of any order is a bicompactness, 3) if a space is ω_α-compact (ω_α is the first number of the might k_α), then it is ϑ_0-compact for $\vartheta_0 < \omega_{\alpha+1}$.

Definition 2 (1). A Banach space E is called regular, if for every $F \in \overline{\overline{E}}$ there exists $x \in E$ such that

$$F(f) = f(x) \text{ for all } f \in \overline{E}.$$

Definition 3. A set L from a regular Banach space E is called ϑ_0-weakly-closed (ϑ_0 is a limiting ordinal number), if together with any sequence $\{x_\vartheta\}_{\vartheta<\vartheta_1}$ (ϑ_1 is a limiting ordinal number, $\vartheta_1 \leqslant \vartheta_0$) to which corresponds a certain x_0 such that

$$\lim_{\vartheta<\vartheta_1} f(x_\vartheta) = f(x_0) \text{ for all } f \in \overline{E}, \qquad (1)$$

it also contains an element x_0.

Definition 4*. By a metric cardinal number of the dimension of a Banach space we define the might of a sequence of elements

$$x_1, x_2, \ldots, x_\vartheta \ldots \quad (\vartheta < \vartheta_0),$$

each of which is at a positive distance from the linear envelope of the preceding and the closed envelope of which coincides with the whole space.

Löwig has shown (2) that this number does not depend on the choice of a sequence of the defined type.

* This definition belongs to Löwig (2).

It can be shown (this has been pointed out by M. Krein together with the author) that the metric cardinal number of a conjugate space $\overline{E}$ is never smaller than the metric cardinal number of the given space E. Hence, in particular, we have the following proposition:

A) *The metric cardinal numbers of the dimension of a regular space E and of its conjugate space coincide.*

II. In connexion with the conception introduced above we prove the following

Theorem 1. *If a set L of a regular space E is bounded and ϑ_0-weakly-closed, then L is ϑ_0-compact in the sense of topology, formed in it by the system of its ϑ_0-weakly-closed subsets* *.

Proof. Let an interval $1 \leqslant \vartheta < \vartheta_0$ have a might k_α. By the remark to Definition 1 it is sufficient to prove that L is ω_α-compact. We shall prove the theorem by the method of transfinite induction. Suppose that the theorem is true for all first numbers ω_γ of the mights $k_\gamma < k_\beta$, where $\beta \leqslant \alpha$. We prove it for the first number ω_β.

Let

$$L_1 \supseteq L_2 \supseteq \ldots \supseteq L_\vartheta \supseteq \ldots; \ (\vartheta < \omega_\beta, \ L_1 \subseteq L) \tag{2}$$

be a sequence of non-empty ω_β-weakly-closed sets. If under the accepted conditions of induction it will be proved that $\prod_{\vartheta<\omega_\beta} L_\vartheta$ is non-empty, then the induction will be performed, and therefore the theorem will be proved. Hence we assume that in (2) everywhere stands the sign $\supset$.

Let

$$x_\vartheta \in L_\vartheta - L_{\vartheta+1} \ (\vartheta < \omega_\beta). \tag{3}$$

Let E_0 be a linear closed envelope $\{x_\vartheta\}_{\vartheta<\omega_\beta}$.

By a theorem of Plessner every closed subspace of a regular space is regular (1). Consequently, E_0 is regular and in virtue of proposition A) there exists a sequence

$$f_1, f_2, f_3, \ldots, f_\vartheta \ldots \ (\vartheta < \omega_\beta) \tag{4}$$

dense in $\overline{E}_0$. We shall denote the smallest closed (in the new topology) set containing $\{x_\vartheta\}_{\vartheta\leqslant\vartheta<\omega_\beta}$ by $L_\vartheta^{(1)}$. Obviously the following is true:

$$L_1^{(1)} \supseteq L_2^{(1)} \supseteq \ldots \supseteq L_\vartheta^{(1)} \supseteq \ldots; \ L_\vartheta \supseteq L_\vartheta^{(1)}; \ (\vartheta < \omega_\beta). \tag{5}$$

We introduce the following abbreviations. If H is a certain set belonging to E and $f \in \overline{E}$, then we denote the set of all different values of a functional $f(x)$ on the elements $x \in H$ by $f(H)$. We use the sign $\prod$ only for the definition of intersection of sets. Besides, the sign $\overset{df}{=}$ is to stand for «equal by definitions».

Let

$$C_\vartheta^{(1)} \overset{df}{=} \prod_{\vartheta\leqslant\vartheta\leqslant\omega} f_1(L_\vartheta^{(1)}) \quad (\vartheta < \omega_\beta). \tag{6}$$

The sets $f_1(L_\vartheta^{(1)})$ are bounded closed sets of numbers [for $L_\vartheta^{(1)}$ are weakly-closed and E is locally weakly-compact (1)], therefore $C_\vartheta^{(1)}$ are non-empty bounded closed sets of numbers.

---

* It can be directly verified that the system of the introduced sets together with the empty set can serve for a system of closed sets.

Let

$$m_{f_1} \overset{df}{=} \sup_{\vartheta<\omega_\beta} \inf C_\vartheta^{(1)} \quad (\inf C_\vartheta^{(1)} \overset{df}{=} \inf_{\alpha\in C_\vartheta^{(1)}} \alpha) \tag{7}$$

From this definition follows the existence of a $\vartheta_1 < \vartheta$ (depending, generally speaking, on f_1) such that

$$\begin{gathered} m_{f_1} = \inf C_\vartheta^{(1)} = \inf C_{\vartheta_1}^{(1)}, \\ \inf C_\vartheta^{(1)} = f_1(y_\vartheta^{(1)}), \text{where } y_\vartheta^{(1)} \in L_\vartheta^{(1)}. \end{gathered} \tag{8}$$

It is easily seen that every linear functional over E is a continuous functional on L in the sense of the introduced topology.

But we have already seen that for $\vartheta > \vartheta_1$ we have $m_{f_1} \in f_1(L_\vartheta^{(1)})$; consequently, the absolute prototype of m_{f_1} in $L_\vartheta^{(1)}$ is a closed non-empty set $L_\vartheta^{(2)}$ for $\vartheta > \vartheta_1$.

Since for $\vartheta < \vartheta_1$ $L_\vartheta^{(1)} \supseteq L_{\vartheta_1}^{(1)}$, the sets $L_\vartheta^{(2)}$ are defined for ϑ satisfying $1 \leqslant \vartheta < \omega_\beta$.

The sets $\{L_\vartheta^{(2)}\}$ are non-empty, closed, and satisfy the following conditions:

$$\begin{gathered} L_1^{(2)} \supseteq L_2^{(2)} \supseteq L_3^{(2)} \supseteq \ldots \supseteq L_\vartheta^{(2)} \supseteq \ldots; L_\vartheta^{(1)} \supseteq L_\vartheta^{(2)}; \\ f_1(L_\vartheta^{(2)}) = m_{f_1}. \end{gathered} \quad (\vartheta < \omega_\beta) \tag{5'}$$

Applying the same method of reasoning, we construct consecutively $f_2(L_\vartheta^{(2)})$, $C_\vartheta^{(2)}$, m_{f_2}, and finally $L_\vartheta^{(3)}$, where $\vartheta < \omega_\beta$. Then the following will be true:

$$C_\vartheta^{(2)} \overset{df}{=} \prod_{\vartheta \leqslant \vartheta < \omega_\beta} f_2(L_\vartheta^{(2)}) \tag{6'}$$

$$\left.\begin{gathered} m_{f_2} \overset{df}{=} \sup_{\vartheta<\omega_\beta} \inf C_\vartheta^{(2)} = \inf C_\vartheta^{(2)} = \inf C_{\vartheta_2}^{(2)}; \quad \vartheta > \vartheta_2, \\ \inf C_\vartheta^{(2)} = f_2(y_\vartheta^{(2)}), \text{ where } y_\vartheta^{(2)} \in L_\vartheta^{(2)}; \quad \begin{matrix}\vartheta_2 \text{ is a certain fixed} \\ \text{ordinal number}\end{matrix} \end{gathered}\right\} \tag{8'}$$

$$\left.\begin{gathered} L_1^{(3)} \supseteq L_2^{(3)} \supseteq \ldots \supseteq L_\vartheta^{(3)} \supseteq \ldots; L_\vartheta^{(2)} \supseteq L_\vartheta^{(3)} \\ f_1(L_\vartheta^{(3)}) = m_{f_1};\ f_2(L_\vartheta^{(3)}) = m_{f_2}. \end{gathered} \quad (\vartheta < \omega_\beta) \right\} \tag{5''}$$

Thus we proceed constructing $L_\vartheta^{(n)}$ till $n < \omega$.

By the assumptions of induction we conclude that $L_\vartheta^{(\omega)} \overset{df}{=} \prod_{n<\omega} L_\vartheta^{(n)}$ is non-empty, if we consider that $L_\vartheta^{(n)}$ are weakly-closed of the order ω, $\omega < \omega_\beta$. Generally $\{L_\vartheta^{(\theta_2)}\}_{\vartheta<\omega_\beta}$ are built as $L_\vartheta^{(2)}$, if θ_0 has a directly preceding number, and are built as $L_\vartheta^{(\omega)}$, if θ_0 is a limiting ordinal number for which $\theta_0 < \omega_\beta$; only it is necessary to take into account that in the case when there exists $\beta - 1$ and $\theta_0 > \omega_{\beta-1}$, the sequence

$$L_\vartheta^{(1)} \supseteq L_\vartheta^{(2)} \supseteq \ldots \supseteq L_\vartheta^{(\theta)} \supseteq \ldots; \ (\theta < \theta_0)$$

has by the assumptions of induction a non-empty intersection (in virtue of the remark to Definition 1).

The sequence $\{L_\vartheta^{(\theta_0)}\}_{\vartheta<\omega_\beta}$ satisfies the following relations:

$$\left.\begin{gathered} L_\vartheta^{(\theta)} \supseteq L_\vartheta^{(\theta_0)}; \\ L_1^{(\theta_0)} \supseteq L_2^{(\theta_0)} \supseteq L_3^{(\theta_0)} \supseteq \ldots \supseteq L_\vartheta^{(\theta_0)} \supseteq \ldots; \ (\theta < \theta_0 < \omega_\beta, \ \vartheta < \omega_\beta) \\ f_\theta(L_\vartheta^{(\theta_0)}) = m_{f_\theta}. \end{gathered}\right\} \tag{5'''}$$

From the relations (5), (5'), (5''), (5''') follows:

$$\left.\begin{array}{c} L_1^{(1)} \supseteq L_2^{(2)} \supseteq \ldots \supseteq L_\vartheta^{(\vartheta)} \supseteq \ldots;\ L_\vartheta^{(1)} \supseteq L_\vartheta^{(\vartheta)};\ (\vartheta < \omega_\beta) \\ f_0(L_\vartheta^{(\vartheta)}) = m_{f\theta};\ (\theta < \vartheta). \end{array}\right\} \tag{9}$$

Hence follows the relation

$$\prod_{\vartheta<\omega_\beta} f_\theta(L_\vartheta^{(\vartheta)}) = m_{f\theta}.\quad (\theta < \omega_\beta) \tag{10}$$

Now let $f \in \overline{E}_0$. Then there exists a sequence of ordinal numbers $\theta_1, \theta_2, \ldots \to \theta^0$ such that

$$f_{\theta_1}, f_{\theta_2}, \ldots \to f.$$

If $L_{\theta^0}^{(\theta^0)}$ contains two points Z_1, Z_2, then from the relation

$$f_{\theta_n}(L_{\theta^0}^{(\theta^0)}) = m_{f_{\theta_n}}$$

follows

$$f_{\theta_n}(Z_1) = f_{\theta_n}(Z_2) = m_{f\theta_n}.$$

Consequently,

$$|f(Z_1) - f_{\theta_n}(Z_2)| \leqslant \|f - f_{\theta_n}\| \cdot \|Z_1\| + |f_{\theta_n}(Z_1) - f_{\theta_n}(Z_2)| \leqslant \|f - f_{\theta_n}\| \cdot R,$$

where R is the radius of a sphere containing L.

In view of the arbitrariness of $Z_1 \in L_{\theta^0}^{(\theta^0)}$, we conclude that the set $f(L_{\theta^0}^{(\theta^0)})$ consists of one number m_f where

$$m_f = \lim_{n\to\infty} m_{f_{\theta_n}}.$$

Thus

$$\prod_{\vartheta<\omega_\beta} f(L_\vartheta^{(\vartheta)}) = m_f.\quad (f \in \overline{E}_0).$$

It is easily seen that $m_f = F(f)$ is a linear functional in $\overline{E}_0$. Consequently, in view of the regularity of E_0, there exists an $x_0 \in E_0$ such that

$$F(f) = f(x_0) \text{ for all } f \in \overline{E}_0.$$

Thus

$$\prod_{\vartheta<\omega_\beta} f(L_\vartheta^{(\vartheta)}) = f(x_0) \text{ for all } f \in \overline{E}_0. \tag{11}$$

Since $L_1^{(1)}$ and x_0 belong to E_0, (11) is also true for all $f \in \overline{E}$.

Let $y_\vartheta \in L_\vartheta^{(\vartheta)}$; then from (11) follows the equality:

$$\lim_{\vartheta<\omega_\beta} f(y_\vartheta) = f(x_0) \text{ for all } f \in \overline{E}.$$

Since $L_\vartheta^{(\vartheta)}$ is ω_β-weakly-closed, we have $x_0 \in L_\vartheta^{(\vartheta)}$ for all $\vartheta < \omega_\beta$.

From the relation (9) follows:

$$x_0 \in \prod_{\vartheta<\omega_\beta} L_\vartheta^{(1)}.$$

From the relation (5) we finally conclude:

$$x_0 \in \prod_{\vartheta<\omega_\beta} L_\vartheta.$$

Thus, by induction, the theorem is proved.

Theoremr 2. *Every regular space is locally bicompact in the sense of topology fo med by the system of all possible sets which are* ϑ*-weakly-closed for any* ϑ *.

This theorem follows directly from Theorem 1 in virtue of the remark to Definition 1.

Theorem 3. *In a regular space E the topology of Tychonoff coincidies with the topologyj introduced in Theorem 2.*

Under the topology of Tychonoff we understand the topology E, gven by neighbourhoods of the form $|f_i(x)-f_i(x_0)|<\varepsilon_i$, $(i=1,2\ldots n)$, where f_i are arbitrary linear functionals, taken in an arbitraty finite number.

Proof. Sets of the form $f(x)\geqslant c$ form a defining system of sets closed in the topology of Tychonoff. The same set $f(x)\geqslant c$ is closed in the topology of Theorem 2. Moreover in the topology of Tychonoff E is a Hausdorff space. Consequently, by Theorem 2, the well-known theorem of the theory of sets (see Hausdorff, Theory of sets, p. 148), by which both topologies coincide, can be applied.

It is interesting to compare Theorems 2 and 3 with the theorem lately obtained by V. Šmulian (3).

«For a Banach space to be regular, it is necessary and sufficient that it should be locally bicompact in the topology of Tychonoff».

We can see that the first (and most difficult) part of this theorem is the direct result of Theorems 2 and 3.

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* Obviously, it is sufficient to demand, instead of a ϑ-weak closure for any ϑ, an ω_α-weak closure, where k_α is the cardinal number of the dimension of the space E.

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ASTRONOMIE

L'ÉTUDE PHOTOMÉTRIQUE DU SYSTÈME U CEPHEI

Par N. I. TCHOUDOVITCHEFF

(Présenté par V. G. Fessenkoff, de l'Académie, le 17. I. 1939)

1. La variabilité de l'éclat U Cephei a été découverte par M-me L. Cerasky le 23 juin 1880.

Déjà les premières observations ont établi l'existence d'une phase constante et l'asymétrie des branches dans le minimum. Plus tard a été découvert le minimum secondaire d'une profondeur pas grande (environ $0^m.1$ dans les rayons visuels et près de $0^m.05$ dans les rayons photographiques).

Des grandes séries des observations photométriques de U Cephei ont été effectuées par M. M. Wendell (1), Dugan (2), Baker (3), Carrosco (4) et Walter (5). M. Dugan a partiellement réussi à expliquer l'asymétrie du minimum principal, en considérant le système comme composé de deux étoiles ellipsoïdales, dont les grands axes sont parallèles et forment avec la droite joignant leur centres de gravité un angle de 23°7.

M. Baker par ses observations confirme les conclusions de Dugan, ainsi que celles de Shapley (6), qui a examiné les résultats des observations de Wendell. Baker considère que les grandeurs relatives de même que les éclats du système sont bien connues.

Mais Walter a établi l'ellipticité différente des composantes, ce qui donnera une forme géométrique toute autre du système.

2. Les observations de U Cephei ont été effectuées par l'auteur avec un astrographe Heide muni d'un objectif Zeiss «Astrotriplet» sur les plaques Ilford Monarch. En tout on a effectué 1125 expositions de deux minutes et demie.

Les éclats des étoiles de comparaison sont déterminés au moyen de comparaison avec N. P. S. (North Polar Sequence). Toutes les plaques sont contrôlées par le microphotomètre photoélectrique de l'observatoire, construit par l'Institut Astronomique à Léningrad et ressemblant le microphotomètre de Schilt. Toutes les observations ont été réunies en 65 points normaux, qui servent à tracer la courbe de l'éclat.

La courbe obtenue possède une asymétrie en minimum principal moindre que les courbes analogues des observations citées dans § 1. Le minimum secondaire s'est trouvé déplacé plus près du minimum principal de la grandeur de l'ordre de $0^d.03$. Chez Wendell et Baker il est déplacé dans le sens opposé de $0^d.10$ environ. Le déplacement du minimum secondaire semble prouver le mouvement de la ligne des apsides dans le système U Cephei.

3. La période. En mesurant la longitude ω du periastre à partir du noeud descendant de l'étoile faible, le moment du minimum sera donné par la formule

$$T_m = T_0 + PE - \frac{Pe}{\pi} \cos(\omega' E + \omega_0), \qquad (1)$$

e étant l'excentricité de l'orbite, ω_0 — la longitude du periastre à l'époque T_0, ω' — la variation de ω pendant une période P.

Carrasco (7) a calculé les différences entre les moments observés des minima et ceux-ci calculés d'après la formule de Wendell

$$T = 2\,407\,890.3007 + 2.492884E.$$

Aux différences $O-C$, trouvées par Carrasco, il faut encore ajouter une valeur correpondante à $E = 8000$ (époque moyenne des observations de l'auteur).

En supposant que la différence $O-C$ comprend: l'erreur de l'époque initiale (ΔT_0), l'erreur de la période (ΔP), adoptée par Wendell, et un membre dépendant du mouvement de la ligne des apsides, on peut écrire:

$$O-C = \Delta T_0 + \Delta PE - \frac{Pe}{\pi} \cos(\omega' E + \omega_0). \qquad (2)$$

La résolution de ces deux équations conduit à la formule suivante:

$$O-C = -0.0050 + 0.0000165\,E - 0.063 \cos(0°.028\,E + 235°) \qquad (3)$$

d'où l'on a

$$e = 0.080.$$

De même, pour le calcul des éphémérides

$$2\,407\,890.2957 + 2.4929005\,E - 0.063 \cos(0°.028\,E + 235°) \qquad (4)$$

et pour la période

$$P = 2.492900 + 0.000031 \sin(0°.028\,E + 235°). \qquad (5)$$

Pour $E = 8000$ on a $P = 2.492932$, ce qui est d'accord avec les résultats de Hellerich (8).

4. La rectification de la courbe de l'éclat a été effectuée en supposant la présence de la pulsation et de la friction des marées. Le meilleur résultat a été obtenu par la rectification qui tenait compte de la friction des marées, au moyen de la formule

$$\Delta m = a + b \sin^{\frac{9}{2}} \frac{\Delta v}{2} + c_0 \cos^2(\theta - \lambda) =$$

$$= a + b \sin^{\frac{9}{2}} \frac{\Delta v}{2} + c' \sin 2\theta + c'' \cos 2\theta.$$

Cela signifie que la pulsation des marées est dans le système de U Cephei pratiquement absente, mais que, par contre, la friction des marées a lieu, ainsi que la rotation des axes qui en résulte (λ).

La table 1 donne le résultat de la rectification, de même que les valeurs de la libration géométrique $\lambda_0 = v - M$ (v désignant l'anomalie vraie, M — l'anomalie moyenne), de la rotation des axes due aux marées ($\lambda_1 = \lambda - \lambda_0$) et de la longitude du periastre.

Table 1

	Courbes visuelles			Courbes photographiques			
	Dugan	Walter	moyenne	Baker	Carrasco	Tchoudovitcheff	moyenne
E	5200	7550	—	5200	7550	8000	—
b	−0.0334	−0.0396	−0.038	−0.0519	−0.0478	−0.0473	−0.049
c_0	+0.0580	+0.0622	+0.060	+0.0544	+0.0538	+0.0420	+0.050
λ	−20°.0	−22°.5	—	−13°.4	−14°	−43°.9	—
λ_0	+8.5	+0.6	—	+8.5	+0.6	−1.4	—
λ_1	−28°.5	−23°.1	−26°	−22°.0	−14°.6	−42°	−26°
ω	+21	+86	—	+21	+86	+99	—
a	+0.0186	+0.0280	—	+0.0231	+0.0240	+0.0007	—

5. La solution a été effectuée pour les courbes de Dugan, de Baker et pour celle de l'auteur. Les courbes ont été rectifiées avec les valeurs moyennes des constantes (Table 1). Quelques raisons font croire que le système de U Cephei contient une étoile moins grande, presque sphérique, brillante et ayant les axes tournés de −26°, et une étoile grande, assez faible, très ellipsoïdale, ayant une libration géométrique (λ_0).

Les formules ordinaires de la détermination des orbites photométriques dans notre cas ne sont pat applicables. La fonction de Russell ψ prend, dans notre cas, la forme:

$$\psi(k'\alpha_n) = \frac{(\sin^2 \Delta v_n - \sin^2 \Delta v_6)\,[1 - Z_1 \cos^2(\theta_9 - \lambda_0)]}{(\sin^2 \Delta v_6 - \sin^2 \Delta v_9)\,[1 - Z_1 \cos^2(\theta_n - \lambda_0)]}. \qquad (6)$$

Ici k' n'est pas encore réel, mais seulement effectif. Sa valeur est de beaucoup plus grande que k réel. Ceci est tout à fait évident, car pendant l'éclipse le grand axe de l'étoile faible est dirigé presque suivant le rayon de vision, et doit figurer dans le rapport des axes après avoir été multiplié par $[1 - Z_1 \cos^2(\theta_n - \lambda_0)]^{\frac{1}{2}}$, autrement dit

$$k = k'\,[1 - Z_1 \cos^2(\theta_n - \lambda_0)]^{\frac{1}{2}}.$$

En se servant de la formule (6) on trouve Z_1 tel que la valeur de k, calculée sur les branches descendante et ascendante de chaque courbe, soit à l'erreur admissible près la même.

Puis, par les formules ordinaires, nous calculons l'inclinaison i et le grand demi-axe de l'étoile a_1', le rayon-vecteur étant pris pour l'unité. Pour obtenir a_1 exprimé en parties du grand demi-axe de l'ellipse, il faut multiplier a_1' par la valeur moyenne du rayon-vecteur pour chaque branche de la courbe, c'est-à-dire

$$a_1 = a_1' r,$$

où r, en se bornant au second degré de l'excentricité, sera donné par

$$r = 1 - e \cos M + e^2 \sin^2 M.$$

Pour calculer l'ellipticité de la composante brillante Z_2 on peut se servir de la formule de Walter (9)

$$\frac{1}{(1-Z)^{\frac{1}{2}}} = \frac{L_1}{(1-Z_1)^{\frac{1}{2}}} + \frac{L_2}{(1-Z_2)^{\frac{1}{2}}}, \qquad (7)$$

en tenant compte des rotations des étoiles dues aux marées et à la libration. Dans notre cas nous aurons, au lieu de (2)

$$\frac{1}{(1-Z\cos^2\lambda)^{\frac{1}{2}}} = \frac{L_1}{(1-Z_1\cos^2\lambda_0)^{\frac{1}{2}}} + \frac{L_2}{(1+Z_2\cos^2\lambda_1)^{\frac{1}{2}}}. \qquad (8)$$

L'ellipticité effective Z est exprimée dans (8) en fonction de c_0 par la relation [10])

$$2Z = c_0 + 0.58b.$$

Table 2

Éléments photométriques du système U Cephei
(l'hypothèse U)

		Photograph.	Visuelles
L'inclinaison de l'orbite	i	84°.2	84°.2
L'excentricité de l'orbite	e	0.080	0.080
La longitude du periastre pour $E = 8000$	ω	99°	99°
Le mouvement de la ligne des apsides pendant une période	ω'	0.028	0.028
Le grand demi-axe de la grande étoile	a_1	0.427	0.414
Le petit demi-axe de la grande étoile	b_1	0.330	0.320
L'ellipticité effective	Z_1	0.4	0.4
L'excentricité de la section équatoriale	E_1	0.636	0.636
Le grand demi-axe de l'étoile moins grande	a_2	0.179	0.173
Le petit demi-axe de l'étoile moins grande	b_2	0.172	0.166
L'ellipticité effective	Z_2	0.08	0.08
L'excentricité de la section équatoriale	E_2	0.285	0.285
Le rapport de rayon des étoiles	k	0.419	0.419
L'éclat de la grande étoile	L_1	0.076 *	0.150
L'éclat de l'étoile moins grande	L_2	0.927	0.854

Conclusion. Les courbes théoriques de l'éclat de Dugan, de Baker et de l'auteur, calculées d'après les données de la table 2, présentent assez bien l'asymétrie du minimum principal, l'asymétrie de la courbe hors de l'eclipse et la position du minimum secondaire. À ce point de vue la courbe de Wendell est moins bonne.

Les 8—10 années à suivre permettront de préciser sensiblement les valeurs indiquées de la période et celles du mouvement de la ligne des apsides, ainsi que vérifier l'hypothèse admise sur le mouvement, car la libration géométrique de l'étoile sombre est maintenant en train de croître et a un autre signe. Cela devra se faire sentir sur l'asymétrie du minimum principal, et sera le plus facile à remarquer, quand le minimum secondaire atteindra son déplacement le plus grand, à savoir, vers l'année 1955.

En tout cas les observations de U Cephei devront être intensifiées, surtout dans la seconde moitié des années 1940, en fixant le minimum secondaire sur les plaques sensibles aux rayons infra-rouges.

Les observations spectroscopiques de cette étoile sont nécessaires à un tel point qu'il est inutile même d'en parler.

L'Institut Astronomique.
Kazan.

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* On a, d'après la courbe de Baker, $L_1 = 0.108$, $L_2 = 0.896$.

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ASTRONOMIE

UNE NOUVELLE ÉTOILE «NAINE BLANCHE»

Par A. DEITCH

(Présenté par V. G. Fessenkoff, de l'Académie, le 13. I. 1939)

En comparant les spectres des étoiles, déterminés à l'observatoire de Bergedorf (1), avec leurs mouvements propres, déduits récemment à l'observatoire de Poulkovo, on est arrivé à découvrir une étoile faible, dont les caractéristiques laissent croire qu'elle appartient à la catégorie peu étudiée des étoiles naines blanches. Ce sont les étoiles d'une densité extrêmement grande et d'un petit éclat absolu. Jusqu'à présent on en connaissait près de 10.

La supposée nouvelle étoile naine blanche a la grandeur photographique visuelle 13.2, le mouvement propre $\mu=0''.288$ par an, dirigé dans le sens opposé au déplacement parallactique ($\psi=358°$). La classe spectrale est *A2p*, la particularité du spectre consistant en raies larges et diffuses, mais en même temps assez intenses de l'hydrogène.

En supposant la grandeur absolue moyenne des étoiles du type *A2* égale à $+2$, la distance de notre étoile sera de l'ordre de 1000 parsecs et la vitesse, d'après le mouvement propre, voisine de 1000 km/sec., ce qui est très invraisemblable.

Il faut croire que l'éclat absolu de l'étoile est considérablement moindre.

La solution définitive du problème, si l'étoile en question appartient à la classe des étoiles naines blanches, ne sera possible qu'après la détermination de sa parallaxe. Les coordonnées de l'étoile sont: $\alpha=9^h\,40^m19^s$ et $\delta=+44°22'\cdot2$ (1900.0) dans l'aire No 29 de Kaptejn, dont la latitude galactique est $+49°$.

Dans la même aire No 29 on a découvert encore une étoile intéressante de grandeur 11.2, dont le mouvement propre est $0''.209$ et dirigé vers l'antiapex ($\psi=174°$). Son spectre est *A5*.

Si l'on considère son mouvement comme parallactique et si l'on en déduit la parallaxe, sa grandeur absolue sera $+9$. Les coordonnées de cette étoile sont: $\alpha=9^h\,43^m10^s$ et $\delta=44°45'.6$ (B. D. $+44°$. 1910).

Dans la zone de Kaptejn $+45$ on a comparé plus de cent mouvements propres ($\geqslant0''.1$ dans un an) des étoiles aux spectres connus; tous les spectres, à l'exception des spectres des deux étoiles citées, sont en majeure partie du type *G* et plus rarement des types *F*, *K* et *M*.

Observatoire Central Astronomique.
Poulkovo.

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ASTRONOMY

ON THE CONSTRUCTION OF A NORMAL SYSTEM OF DECLINATIONS ON THE BASE OF ZENITH SYMMETRY OF SYSTEMS DERIVED FROM OBSERVATIONS

By V. G. SHAPOSHNIKOV

(Communicated by V. G. Fessenkov, Member of the Academy, 13. I. 1939)

There exists a possibility of constructing with a high degree of accuracy the normal system of declinations for both equatorial and all other declinations. The absolute errors of individual systems are of little importance for this method which has not been used till now.

The principles of the method are—the symmetry of refraction and the symmetry of systematical errors of observations of zenith distances, the zenith of the observatory being taken as the centre of symmetry. Provided that both these principles are fulfilled, we have the following property of an individual system: if the arcs of the meridian symmetrically situated on both sides of the zenith are equal in an individual system of declinations of a given observatory, then these arcs are actually equal. This statement is valid for points in lower culmination, provided that a special system of declinations is formed for such points. This special system is to be formed on the base of observations of zenith distances in lower culmination, provided that the elements of reduction (latitude and refraction) are the same for the whole sky.

The symmetry of refraction is satisfied with considerable accuracy, though it would be difficult to evaluate it exactly. The symmetry of systematical errors is one of the most natural and attainable conditions to be fulfilled in meridian observations. Moreover, the possible deviations from the symmetry of systematical errors for a number of observatories are likely to be of a casual character.

For the construction of a normal catalogue we have recourse to a system which is to be improved. We may assume our preliminary system to be the first approximation of the normal system. Among the systems constituting the material for the construction of a normal catalogue, there must be a sufficient number of systems including safely determined declinations near the poles, i. e. of systems in which the latitude has been derived from a sufficiently great number of observations of upper and lower culminations of stars near the pole. In case latitude is determined in another way, the system does nevertheless not lose its validity to establish a normal system of all the declinations except for stars in the vicinity of the pole. Full weight would have catalogues of low latitudes which contain no high declinations at all. It may be assumed that the preliminary system for the two zones in the

pole's vicinity is the normal system and there is no need to improve it in these zones.

There are cases in which the problem may be solved in some sense by means of two catalogues formed at observatories of both hemispheres whose latitudes are definitely chosen. The simplest case is that of latitudes $+30°$ and $-30°$. The declinations $+30°$ and $-30°$ may then be determined, as is easily seen on the basis of the general principle of the zenith symmetry or by means of equations of form (2). It is supposed further that we dispose of a number of catalogues formed at different latitudes of both hemispheres.

Let us determine the points of division of the semicircumference, say, into 12 parts, thus obtaining points corresponding to the declinations at intervals of 15°:

$$+75°,\ +60°,\ +45°,\ +30°,\ +15°,\ 0°,\ -15°,\ -30°,\ -45°,\ -60°,\ -75°.$$

The interval between the points of division is not arbitrary. The conditions to be satisfied when selecting it, will be given further.

Let us put

$$\Delta\delta_{+75},\ \Delta\delta_{+60},\ \ldots,\ \Delta\delta_{0},\ \ldots,\ \Delta\delta_{-60},\ \Delta\delta_{-75} \qquad (1)$$

the sought for systematical corrections of the preliminary system for the formation of a normal system. The quantities $\Delta\delta^i_\delta$ are the differences: the individual system i minus the preliminary system. The values $\Delta\delta^i_{+90}$ and $\Delta\delta^i_{-90}$ are not necessarily null. Suppose we have a catalogue i, formed at the latitude φ_i. Considering the arc from the zenith to the zenith distance z, we have the zenith symmetry condition ($z < 90° - |\varphi_i|$) in the form:

$$\Delta\delta_{\varphi_i - z} - 2\Delta\delta_{\varphi_i} + \Delta\delta_{\varphi_i + z} = \Delta\delta^i_{\varphi_i - z} - 2\Delta\delta^i_{\varphi_i} + \Delta\delta^i_{\varphi_i + z}. \qquad (2)$$

For $z > 90° - |\varphi_i|$ we have:

$$\Delta\delta_{\varphi_i - z} - 2\Delta\delta_{\varphi_i} - \Delta\delta_{180 - \varphi_i - z} = \Delta\delta^i_{\varphi_i - z} - 2\Delta\delta^i_{\varphi_i} - \Delta\delta^{i*}_{180 - \varphi_i - z};\ \varphi_i > 0 \qquad (3N)$$

and

$$\Delta\delta_{\varphi_i + z} - 2\Delta\delta_{\varphi_i} - \Delta\delta_{z - \varphi_i - 180} = \Delta\delta^i_{\varphi_i + z} - 2\Delta\delta^i_{\varphi_i} - \Delta\delta^{i*}_{z - \varphi_i - 180};\ \varphi_i < 0 \qquad (3S)$$

where by * are marked the systematical corrections of declinations of a special system, formed by means of observations of the lower culminations. The presence of equations of the form (3) would contribute to the accuracy of the result, yet they are not indispensable. We assume the possibility of linear interpolation of corrections of the preliminary system between the cardinal points of division selected by us. This condition being fulfilled, the left parts of the equations of the form (2) and (3) are readily transformed for arbitrary φ_i and z introducing the primary unknowns (1).

The weight of an equation of the form (2) or (3) would be nearly constant for any z, provided that the stars observed are always arranged in pairs symmetrically on both sides of the zenith. If this condition is not fulfilled, the weight would decrease with the increase of z.

A case of the strictly sufficient number of equations is characteristic of the accuracy attainable in the results. It is the case of three catalogues, observed in latitudes $+30°$, $-30°$ and $+45°$. We have 11 equations of the form (2) of 11 unknowns. Putting the mean error of an equation $\pm\varepsilon$, we obtain the following mean errors of unknowns:

$$\varepsilon_{+75} = 1.9\varepsilon;\quad \varepsilon_{+30} = 0.7\varepsilon;\quad \varepsilon_{-15} = 2.5\varepsilon;\quad \varepsilon_{-60} = 1.6\varepsilon;$$
$$\varepsilon_{+60} = 0.9\varepsilon;\quad \varepsilon_{+15} = 1.4\varepsilon;\quad \varepsilon_{-30} = 0.7\varepsilon;\quad \varepsilon_{-75} = 1.7\varepsilon.$$
$$\varepsilon_{+45} = 0.7\varepsilon;\quad \varepsilon_{0} = 1.4\varepsilon;\quad \varepsilon_{-45} = 2.5\varepsilon;$$

The accuracy of the results is satisfactory, taking into account that s depends only on the deficiency of zenith symmetry of the catalogues considered (except the errors in determination of systematical differences of the catalogues).

The equations of condition for different z in the same catalogue are not quite independent. One should, therefore, not look for the corrections of the preliminary system at too small intervals. To obtain a sufficient accuracy in a linear interpolation of the corrections of a preliminary system the interval should not be too large either.

If the catalogues cover a considerable period of time, it is possible to obtain proper motions in declination by introducing the corrections of proper motions of a preliminary system as unknowns into conditions of form (2) and (3).

If individual catalogues fulfil the condition of zenith symmetry in separate zones of right ascension, no errors of form $\Delta\delta_\alpha$ are present.

Data of high accuracy for controlling the normal system of declinations may be obtained by means of observation with zenith-telescopes. If we put the declinations of the stars observed s, ζ, and n (ζ nearly equal φ), we get:

$$\Delta\delta_s - 2\Delta\delta_\zeta \pm \Delta\delta_n = 2(\varphi_\zeta - \varphi_{n,s}), \qquad (4)$$

where $\Delta\delta$ are as before systematical corrections of the preliminary system; φ_ζ, latitude calculated from observations of a star near the zenith; $\varphi_{n,s}$, latitude calculated from observations of a pair of stars with declinations n and s, the declinations for calculation of latitudes being taken from the preliminary system. Signs $\pm$ refer to the upper and lower culminations respectively. Four stations of such a kind in different latitudes would be sufficient, two in the northern hemisphere and two in the southern one. The stars in the vicinity of the poles must be determined in the usual way by means of the meridian or vertical circles. The realization of such a scheme would result in a high precision system of declinations for the epoch of observations. If this work is repeated in half a century or so, a system of proper motions of high degree of accuracy will be established.

It would be very desirable for such a work to dispose of instruments for measuring differences of zenith distances of the order 2—3°. This would enable us to introduce the least possible alterations into the fundamental lists of stars.

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PHYSICS

USE OF FLAME IN SPECTRUM ANALYSIS

By S. L. MANDELSTAM

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1. In the application of a flame as an excitation source for spectrum analysis work the substance under investigation is introduced into the flame in the form of the salts of the metal on hand, such as chlorides, sulphides, etc. Outstanding interest belongs to the question of how the dissociation process of these salts runs its course in a flame, as well as whether or not there exists a dependence between the intensity of the lines of the element analyzed and the kind of compounds in which it has been introduced into the flame. Closely related thereto is also the question of the influence of the solutions of salts of some other elements upon the intensity of the lines of the element analyzed.

The experimental material mentioned in the literature on the subject, both dealing with spectrum analysis and devoted to the electrical conductivity of a flame, bears a very contradictory character. A number of investigators failed to observe any such influence, whereas others pointed out in their works that an influence of such a kind did definitely exist ([1,2]).

To the number of the latter belongs the work performed lately by T. F. Borovick-Romanova ([3]). In this work is to be found an investigation into the influence of large amounts (0.1—4%) of NaCl, KCl, NaBr, NaF, Na_2SO_4 and other compounds upon the intensity of the lines of rubidium introduced in solution in the form of RbCl (0.001% by weight). It was discovered that the existence of NaCl, NaBr, NaI in solution lowers in a great measure the intensity of the lines of Rb; that KCl exerts a weakening effect, though less pronounced than NaCl; while the presence of NaF and Na_2SO_4 has no bearing upon the intensity of the lines of Rb; thus, the existence of a certain relationship between the strength of the molecules and the magnitude of the influence has been established.

2. It may at present be considered as definitely established that excitation in a flame bears in the main a thermal character; the surmise that the dissociation of molecules introduced into the flame is also thermal should also be regarded as sufficiently probable.

Taking our starting point in these considerations, we have performed in the present work the calculations of the magnitudes of dissociation for those compounds and conditions of work, under which the investigations of Borovick-Romanova were accomplished. At the bottom of our computation lies the approximate formula for the calculation of the equilibrium constant:

$$\lg K_p = -\frac{W_p}{4.57T} + \sum \nu \cdot 1.75 \lg T + \sum C_i. \qquad (1)$$

Calculation according to more exact formulae was held to be less expedient, since the other data had not been determined with a sufficient degree of accuracy. The great variety of compounds available in the flame, the diversity of the temperature of the flame, the different rate of diffusion of various components, etc., in fact all these factors together, compelled us to confine ourselves to obtaining but very approximate results, which could rather be qualified as qualitative than quantitative.

In formula (1) K_p denotes the equilibrium constant with constant pressure; W_p, the heat of formation of molecules; $\sum v$, the change in the number of moles of gas in the reaction; C_i are conventional chemical constants; T is absolute temperature, assumed by us in the subsequent to be equal to 2000°. (The work of Borovick-Romanova was effected with the help of a burner fed by a mixture of air and acetone vapours whose temperature was, to all appearances, very close to 2000°.) The data for W_p were taken by us from Mayer and Helmholz's works (4). The values of K_p calculated according to these data with the help of formula (1) are represented in Table 1.

Table 1

Compound (in gaseous state)	Reaction (both comp. are gaseous)	Heat of formation, W_p	K_p (T=2000°)
RbCl	Rb + Cl	97 400	$6.3 \cdot 10^{-6}$
RbBr	Rb + Br	90 500	$4 \cdot 10^{-5}$
RbF	Rb + F	130 600	$4 \cdot 10^{-9}$
NaCl	Na + Cl	97 700	$6.3 \cdot 10^{-6}$
NaBr	Na + Br	86 000	$1.1 \cdot 10^{-4}$
NaF	Na + F	130 000	$4 \cdot 10^{-9}$
KCl	K + Cl	101 400	$2.5 \cdot 10^{-6}$

For the equilibrium state of dissociation of RbCl in the absence of admixtures the following relations are available:

$$\frac{p_{Rb}\, p_{Cl}}{p_{RbCl}} = 6.3 \cdot 10^{-6}; \qquad \begin{matrix} p_{Rb} + p_{RbCl} = P_{RbCl} \\ p_{Cl} + p_{RbCl} = P_{RbCl} \end{matrix} \qquad (2)$$

where P_{RbCl} is partial pressure of RbCl before the dissociation, not exactly known. In the subsequent we shall give all our calculations for three values of P_{RbCl}: 10^{-5}, 10^{-6}, 10^{-7} atm. The pressure available in Borovick-Romanova's work was apparently close to 10^{-6}. We proceed here from the evaluation found in the literature of the visibility boundary of the spectrum of rubidium in flame corresponding to a concentration of $2 \cdot 10^{13}$ mol/cm³ (2) and determined from the amount of the solution admitted into the flame per unit of time and the rate of the blowing of the gas.

From relations (2) we then obtain the following values for the partial pressures of the atoms of Rb and for the respective degree of dissociation.

Table 2

P_{RbCl}	p_{Rb}	Degree of dissociation, %
10^{-5}	$0.55 \cdot 10^{-5}$	55
10^{-6}	$0.85 \cdot 10^{-6}$	85
10^{-7}	$0.98 \cdot 10^{-7}$	98

With the availability of large amounts of NaCl in the solution, the chlorine, liberated in result of dissociation, shifts the degree of dissociation of RbCl, reducing the partial pressure and consequently the intensity of the

spectrum lines of rubidium as well. In this case the joint solution of the following relations* will be necessary:

$$\left.\begin{aligned} &\frac{p_{Rb}\,p_{Cl}}{p_{RbCl}} = 6.3\cdot 10^{-6}; \\ &\frac{p_{Na}\,p_{Cl}}{p_{NaCl}} = 6.3\cdot 10^{-6}; \end{aligned}\right| \quad \left.\begin{aligned} &p_{Rb} + p_{RbCl} = P_{RbCl} \\ &p_{Na} + p_{NaCl} = P_{NaCl} \\ &p_{Cl} + p_{RbCl} + p_{NaCl} = P_{RbCl} + P_{NaCl}. \end{aligned}\right| \qquad (3)$$

Effecting all the calculations for a 1% concentration of sodium in the solution, i. e. assuming that $P_{NaCl} = P_{RbCl}\cdot 2\cdot 10^3$ (due account is taken of the difference in the molecular weights of NaCl and RbCl), the following magnitude is obtained for the partial pressure of the atoms of Rb (this pressure in the case of the absence of NaCl is taken for 100), three pressures being taken for P_{RbCl} (Table 3).

Table 3

P_{RbCl}	0.001% RbCl	0.001% RbCl + 1% NaCl	0.001% RbCl + 1% KCl	0.001% RbCl + 1% NaBr	0.001% RbCl + 1% NaF
10^{-5}	100	3.5	6.5	6.5	~100
10^{-6}	100	7.5	13	13	~100
10^{-7}	100	18	25	33	~100

i. e. for the case of $P_{RbCl} = 10^{-6}$, the introduction of NaCl reduces the mean concentration of the atoms of Rb and, consequently, also the intensity of the lines of Rb nearly 13-fold.

3. Calculations for the case of the presence of KCl are performed in an altogether analogous way. Their results are to be found summed up in Table 3. However, the weakening effect obtained according to these calculations is, to all appearances, more considerable than that observed in the experiments of Borovick-Romanova. Possibly, we deal here with a certain compensation for the lowering of the intensity of Rb-lines, brought about by a shift in the degree of dissociation at the expense of the stimulation of emission, due to collisions of the second kind with the excited atoms of K. The resonance levels of both elements according to their energies are very close to each other (0.01 V) and, consequently, the effect for these levels may be quite considerable, as this has been repeatedly observed for similar cases [5].

In the evaluation of the influence of NaBr the following system of relations has to be jointly solved:

$$\left.\begin{aligned} &\frac{p_{Rb}\,p_{Cl}}{p_{RbCl}} = 6.3\cdot 10^{-6}; \\ &\frac{p_{Rb}\,p_{Br}}{p_{RbBr}} = 4\cdot 10^{-5}; \\ &\frac{p_{Na}\,p_{Br}}{p_{NaBr}} = 1.1\cdot 10^{-4}; \\ &\frac{p_{Na}\,p_{Cl}}{p_{NaCl}} = 6.3\cdot 10^{-6}; \end{aligned}\right| \quad \left.\begin{aligned} &p_{Rb} + p_{RbCl} + p_{RbBr} = P_{RbCl}; \\ &p_{Cl} + p_{RbCl} + p_{NaCl} = P_{RbCl}; \\ &p_{Na} + p_{NaBr} + p_{NaCl} = P_{NaBr}; \\ &p_{Br} + p_{NaBr} + p_{RbBr} = P_{NaBr}. \end{aligned}\right|$$

* Here, as well as in the subsequent, the existence of the compounds of Cl_2, Br_2, F_2 is not accounted for, since, due to the small strength of their molecules, they are fully dissociated.

The obtained results are summed up in Table 3.

In considering the influence of NaF, an analogous system has to be solved. Due account should, however, be taken here of the very strong bonds between fluorine and carbon, as well as its oxides and hydrogen which practically completely bind the fluorine, liberated in result of the dissociation of NaF. Thus, if the reaction $H_2O_{gas} + 2F = 2HF + \frac{1}{2}O_2 + 134.4$ is properly accounted for and if the value 10^{-2}—10^{-1} atm. is assumed for the partial pressure of water vapour produced in the combustion of acetone, then there will be practically no influence of NaF upon the intensity of the Rb lines (Table 3).

4. We shall confine ourselves in the present to the given example. We believe that, notwithstanding the but semi-quantitative character of the computations, the experimental picture observed has received a perfectly convincing explanation.

The above-mentioned calculations also permit, in our opinion, to account for the contradictory observations of different authors. The salts usually applied are on the very boundary of full dissociation and, consequently, very slight changes in experimental procedure—for instance, the use in certain runs of a hotter flame (say, an air-acetonic flame with a T of 2500°)—may involve so complete a dissociation that the influence of the composition of the compounds and the admixture will practically not be observable.

5. It would seem that the above-said reasonings may also be extended to the arc and the spark. As a rule, owing to considerably higher temperature, we should expect here a total absence of the aforesaid influences. Yet, in certain particular cases these influences were, in point of fact, observed. Thus, for instance, Neidler ([6]), while working with a copper arc, observed a difference in the intensities of the lines of W, Zr and other elements according to the kind of compound. This was precisely in the direction which could be expected taking into account the data on the strength of bonds in the compounds in question. With a carbon arc, i. e. with a higher temperature, the influence of the kind of compound was, as could be expected, liable to decline. It is possible that the specific way of inroducing the material into the arc (strewing on) could be held responsible for the case at hand. Gerlach and Ruthardt ([7]) have in a similar way observed in a spark the dependence of the intensities of the lines of certain metals upon the kind of compounds. A decrease in intensity was noticed to run in the following way:

$$MeNO_3 > MeI > MeBr > MeCl > Me_2SO_4.$$

It is precisely in this direction that the increase of the strength of molecules runs.

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GEOPHYSICS

GENERAL THEORY OF A VISIBILITY-METER

By W. W. SCHARONOW

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In the present paper, the term «visibility-meter» is applied to such a device in which, by means of artificial turbidity located in the line of vision of the observer, the visibility of a certain contrast is being lessened till its complete disappearance. Devices recommended by Dannmeyer (1), Wigand (2), Bennett (3), Jones (4) and Scharonow (5) may serve as an example of such apparatuses. In literature one may come across such opinion that the devices of this kind give but subjective characteristics of the visibility and thus cannot serve for the quantitative measurement of the turbidity of the atmosphere. In order to confirm this opinion, there are usually put forth failures which followed all attempts to apply Wigand's visibility-meters in practical work. In the paper published by the writer in 1933 (6) it was shown that the real cause of these failures lies in the wrong theory suggested by Wigand himself as regards his devices, and that the true theory of a visibility-meter must proceed from the photometric theory of contrasts and the Weber-Fechner law. Essays to build up such a theory as regards individual types of visibility-meters and particular cases of observation have been made by different authors (7). In the present paper the writer gives the principles of the general theory of a visibility-meter in its application to the determination of both the extinction factor of the air in the horizontal direction and the range of visibility of an arbitrary object.

The optic action of the turbid medium of a visibility-meter consists in the fact that the brightness B of the object observed with the aid of the device is being decreased to a certain value $TB\,(T<1)$ and that, at the same time, there is the superposition upon it of the brightness b, this brightness being that of the light scattered by the medium; thus, the brightness B', visible through the device, will be as follows:

$$B' = BT + b. \qquad (1)$$

The contrast formed by two brightnesses, B_1 and B_2, when observed through the turbid medium of the device, will prove to be decreased. In the visibility-meter, the scale-reading x corresponds to the limit of visibility of the contrast, and is determined by the condition, which states that the value of the contrast formed by two brightnesses, B_1' and

B_2', should be equal to the threshold of the contrast sensitivity of the eye ε:

$$\frac{B_1' - B_2'}{B_1'} = \frac{B_1 - B_2}{B_1 + \frac{b}{T}} = \varepsilon. \qquad (2)$$

In further lines, the ratio $\frac{b}{T}$ will be termed as «turbidity» and lettered with M; it is supposed that M is the monomial function of the scale-reading x and the form of this function is known to the observer:

$$M = \frac{b}{T} = f(x). \qquad (3)$$

It should be pointed out that in some visibility-meters the increase of x is followed also by a certain diffusion of contours of the object under observation; this results in a variation in the value of ε. The theory set forth in the present paper does not extend to all these cases; for the future, it should be assumed that the working principle of the device and the conditions of observation are such that the threshold of the contrast sensitivity for the given group of observations can be considered as a constant one.

In the practice of observation of the daylight visibility, one has to deal with contrasts of the following types:

I. Surfaces forming the contrast belong to two terrestrial objects and have between them and the observer an equal distance l. The brightnesses B_1 and B_2 of these surfaces are different owing either to different reflexion factor or to different illumination.

II. Surfaces forming the contrast belong to two terrestrial objects at different distances, l_1 and $l_2 = l_1 + \Delta$, from the observer. The true brightnesses of these surfaces can be either same or different.

III. The contrast is formed by a terrestrial object (brightness B) and the sky at the horizon (brightness B_H).

Optical conditions of the layer of the air between the observer and the object can be characterized by the following known values:

$$\tau = 10^{-al}, \qquad \beta = B_H(1 - 10^{-al}), \quad \mu = \frac{\beta}{\tau} = B_H(10^{al} - 1), \qquad (4)$$

where τ is transmission factor; a — extinction factor; l — distance to the object; β — brightness of light scattered by the air layer in some direction; B_H — brightness of sky in the same direction; as regards the value μ, it will be termed as «turbidity» of the given layer of the air.

Let the contrast of the type I be observed twice with the aid of a visibility-meter. In the course of the first observation, its distance from the observer is so small that one may reckon $\tau = 1$ and $\beta = \mu = 0$. In this case, the turbidity M_0, corresponding to its disappearance in the visibility-meter, can be directly found from the equation (2):

$$M_0 = \frac{1}{\varepsilon}[B_1(1 - \varepsilon) - B_2]. \qquad (5)$$

Let the second observation be carried out from a rather great distance l. In this case the brightnesses B_1'' and B_2'' of objects forming the contrast are to be determined by the joint action of turbidities of both the atmosphere and the device:

$$B_1'' = (B_1\tau + \beta)T + b; \quad B_2'' = (B_2\tau + \beta)T + b; \qquad (6)$$

as regards the condition of disappearance in the visibility-meter, it is determined by the following equation:

$$\frac{B_1' - B_2''}{B_1''} = \frac{(B_1 - B_2)\tau}{B_1\tau + \beta + M} = \varepsilon. \qquad (7)$$

Hence

$$M = \frac{\tau}{\varepsilon}[B_1(1-\varepsilon) - B_2] - \beta. \qquad (8)$$

In using the equation (5), the general equation of the theory of a visibility-meter can be written as follows:

$$M = M_0\tau - \beta. \qquad (9)$$

It becomes quite obvious that this equation is true for contrasts of other types as well. And in fact, if in the contrast of the type II the brightness of a more remote object will be designated as B_2, its brightness B_2' when observed from the distance $l_2 = l_1 + \Delta$ will give:

$$B_2' = B_2 10^{-al_2} + B_H(1 - 10^{-al_2}) = B_2'' 10^{-al_1} + B_H(1 - 10^{-al_1}),$$

where the symbol B_2'' shows the brightness of this object, as it is seen from the object which lies nearer:

$$B_2'' = B_2 10^{-a\Delta} + B_H(1 - 10^{-a\Delta}).$$

From this it is apparent that equations (6), (7) and (8) can be applied to the contrast of the type II, when B_2'' is being substituted for B_2 and the distance calculated from the object which lies nearer. The same formulae can be applied to contrast of the type III as well, if one reckons $B_1 = B_H$ and $B_2 = B$. This is obtainable in virtue of the identity:

$$B_H' = B_H\tau + \beta = B_H.$$

Fundamental equation (9) can be utilized for the determination of the photometric characteristics of the atmosphere in the given direction B_H and a. It is evident that this requires the performance of 3 observations of a certain contrast: the first of them is to be carried out from a close distance, the second and the third — from two far distances l_1 and l_2; in result, one will obtain turbidities M_0, M_1 and M_2. Thus, following expressions can be written:

$$1 - \frac{M_1}{M_0} = \left(1 + \frac{B_H}{M_0}\right)(1 - 10^{-al_1}); \quad 1 - \frac{M_2}{M_0} = \left(1 + \frac{B_H}{M_0}\right)(1 - 10^{-al_2}).$$

Hence

$$\frac{1 - \frac{M_2}{M_0}}{1 - \frac{M_1}{M_0}} = \frac{1 - 10^{-al_2}}{1 - 10^{-al_1}} = \gamma. \qquad (10)$$

Following symbols are brought in: $\frac{l_2}{l_1} = n$; $10^{-al_1} = \tau$; $10^{-al_2} = \tau^n$.

Then in order to find out τ, the following equation is obtained:

$$\tau^n - \gamma\tau + \gamma - 1 = 0. \qquad (11)$$

This equation, being solved with the aid either of tables or nomograms, leads to the possibility to determine the coefficient of extinction a. The latter being known, the second characteristic of turbidity is easily found out in the following form:

$$\frac{B_H}{M_0} = \frac{\tau_1 - \frac{M_1}{M_0}}{1 - \tau_1}. \qquad (12)$$

If the problem of measurements consists only in the determination of the value of a, simplified methods of observation can be applied. Thus, when a black object ($B=0$) against the sky is observed first at a close distance, and then at a far distance l, provided that the value ε compared to τ is to be neglected, following very simple equations are obtained:

$$\tau=\frac{M}{M_0}; \quad a=\frac{1}{l}\log\frac{M_0}{M}. \tag{13}$$

The range of visibility L of a certain object represents obviously that distance, at which the contrast B_1, B_2 reaches the threshold of the contrast sensitivity under the observation without the visibility-meter. Hence

$$\frac{B_1-B_2}{B_1+\mu}=\varepsilon. \tag{14}$$

When comparing this equation to equation (2), we obtain:

$$\mu=B_H(10^{aL}-1)=M_0, \tag{15}$$

wherefrom L is found out:

$$L=\frac{1}{a}\log\left(\frac{M_0}{B_H}+1\right). \tag{16}$$

This formula, moreover, is true for all the three types of contrasts. The value $\frac{M_0}{B_H}$ enclosed in it can be determined apart from the equation (12), by the following method as well. A black screen ($B=0$) is placed in the azimuth of the object and then observed against the sky from a close distance. The turbidity M_4, obtained under such an observation, will evidently be as follows:

$$M_4=B_H\frac{1-\varepsilon}{\varepsilon}. \tag{17}$$

The observation, from a close distance, of the contrast, the range of visibility of which is to be determined, gives a turbidity designated as M_0; thus, it can be written:

$$\frac{M_0}{B_H}=\frac{M_0}{M_4}\,\frac{1-\varepsilon}{\varepsilon}. \tag{18}$$

It should be noted that for the computation of the range of visibility according to the known value a, it is of greatest use to have, for different objects under different conditions of weather and illumination, the tables of the following value:

$$\frac{M_0}{B_H}=\frac{B_1(1-\varepsilon)-B_2}{\varepsilon B_H}. \tag{19}$$

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GEOPHYSICS

TEMPERATURE WAVES IN THE MONSOON FIELD

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It is known that the difference of temperatures of the air above the sea and the continent forms a monsoon field over them. This field has two currents superposed: colder air moves in the lower stratum several hundred meters thick; warmer air occupies the higher stratum. An accurate solution of unstationary problems connected with the monsoon field is most probably impossible. Therefore we shall try to solve one of such problems applied to some model of a monsoon field, without any attempts to accuracy.

Suppose that the vertical distribution of the wind velocities is determined by some law somewhat resembling that found in one of our previous works (1). Instead of the velocities at various altitudes of the lower monsoon layer, we shall introduce a resulting velocity U, giving a current in this layer similar both in size and direction. Suppose that the current in the upper layer is determined by the resulting velocity equal to U in size and having the opposite direction. We shall examine a certain small disturbance arisen in the monsoon field. Let it be, for instance, a small rise of mean temperature of the whole monsoon layer to θ in the given section of the field. Then the atmospheric pressure in this section will be changed to a certain value p. Consequently, the resulting velocity U (and the corresponding $-U$) will be changed too, having got an unstationary increase which, generally speaking, will give the parallel component (u) and the meridian one (v). Let the singular vector of acceleration form a certain angle α with the pressure gradient (grad p): the angle caused by the action of Coriolis force on the moving air masses.

If $\cos\alpha = m$ and $\sin\alpha = n$ and if the air density is equal to δ, the acceleration components along the co-ordinate axes will be expressed by

$$\frac{\partial u}{\partial t} = -\frac{m}{\delta}\frac{\partial p}{\partial x} - \frac{n}{\delta}\frac{\partial p}{\partial y}, \qquad (1)$$

$$\frac{\partial v}{\partial t} = -\frac{m}{\delta}\frac{\partial p}{\partial y} + \frac{n}{\delta}\frac{\partial p}{\partial x}. \qquad (2)$$

But the derivatives of p may be expressed in their turn by the derivatives of θ, using one of the basic equations of the monsoon theory (2):

$$\text{grad } p = -\Pi \text{ grad } \tau. \qquad (A)$$

In this equations, found in the paper cited under numbers (19) and (73), T and τ respectively denote: the mean temperatuse of all the active

(«two-layer») monsoon stratum (in the investigations of the theoretic model in the simplest case), or the temperature anomalies (in the examinations of climatological charts). Π denotes a constant, the dimension of which is deduced from the formula (A) itself. Thus, according to (1), (2) and (A):

$$\frac{\partial u}{\partial t} = \frac{\Pi}{\delta}\left(m\frac{\partial \theta}{\partial x} + n\frac{\partial \theta}{\partial y}\right), \tag{3}$$

$$\frac{\partial v}{\partial t} = \frac{\Pi}{\delta}\left(m\frac{\partial \theta}{\partial y} - n\frac{\partial \theta}{\partial x}\right). \tag{4}$$

The change of the velocities immediately causes the change of the current of heat energy transferred from the sea to the continent (or in the opposite direction) by the air masses; and this, in its turn, is to influence the change of the value θ of the original thermal disturbance.

The whole process can as yet be described but approximately; we must take a certain mean difference of temperatures Θ between the upper and lower monsoon layers, and suppose that Θ is slightly changed with the changes of x and y.

Namely, as it is easy to show, neglecting the infinitesimals of second order:

$$\frac{\partial \theta}{\partial t} = \Theta\left[\frac{\partial u}{\partial x} + \frac{\partial v}{\partial y}\right]. \tag{5}$$

Let us differentiate (5) to t and find for the compound derivatives for x, t and for y, t, which we have got in the right part, the expressions got by differentiation of (3) to x and of (4) to y. As a result, we shall get the simple equation:

$$\frac{\partial^2 \theta}{\partial t^2} = m\frac{\Pi\Theta}{\delta}\left[\frac{\partial^2 \theta}{\partial x^2} + \frac{\partial^2 \theta}{\partial y^2}\right]. \tag{6}$$

This equation describes a temperature wave: either freely progressive with the velocity

$$c = \sqrt{m\frac{\Pi\Theta}{\delta}}, \tag{7}$$

or a standing one under corresponding limit conditions.

Let us consider the behaviour of waves inside a certain limited nodal line. Let this line, at which temperature is stationary, be a circle with the radius ρ. If ρ is not too large, it is possible to neglect the incurvation of the Earth's surface and to apply the equation (6) got for a plane.

For convenience let us transform (6) to the polar system of co-ordinates (r, ψ) with the pole in the centre of the nodal circumference. It must be also taken into consideration that in any given point of the field θ must be connected both with t and ψ by certain functions allowing the analysis into Fourier series.

Then the differential equation of the waves will assume the form:

$$\frac{\partial^2 \theta}{\partial r^2} + \frac{1}{r}\frac{\partial \theta}{\partial r} + \left(k^2 - \frac{s^2}{r^2}\right) = 0. \tag{8}$$

This equation is integrated in Bessel functions, with limit condition

$$|\theta|_{r=\rho} = 0. \tag{9}$$

We shall need its integral:

$$\theta = AJ_1(kr)\cos(\sigma t + \psi + \varepsilon), \tag{10}$$

in the expression of which on the bases of (9)

$$J_1(k\rho) = 0. \tag{11}$$

Thus, in order to find the maximum period of the proper oscillations we can find in the tables

$$k\rho = 3.83. \tag{12}$$

As, on the other hand,

$$T = \frac{2\pi}{kc}, \tag{13}$$

on the ground of (13), (7) and (12)

$$T = \frac{2\pi}{3.83}\cdot\frac{\rho}{c} = 1.64\rho\sqrt{\frac{\delta}{m\Pi\Theta}}. \tag{14}$$

Here the value m remained indetermined. Let us try to find an approximate value for it. We must take into consideration that the problem just solved is perfectly similar from the formal point of view to that of seiches in a small sea of constant depth and circular form, i. e. to the problem solved by Lamb [3] (as a development of Kelvin's ideas). But in this case the displacements ξ and η of the air particles respectively along r and normally to it (in the horizontal plane) may be expressed with θ by the equations resembling Lamb's equations for the horizontal displacements of the water particles, where these displacements are given as functions of the rise of the sea surface ζ. Namely, we are able to write:

$$\xi = M\left(\frac{\partial\theta}{\partial r} - \frac{2i\overline{\omega}}{\sigma r}\frac{\partial\theta}{\partial\psi}\right), \tag{15}$$

$$\eta = iM\left(\frac{2\overline{\omega}}{\sigma}\frac{\partial\theta}{\partial r} - \frac{i}{r}\frac{\partial\theta}{\partial\psi}\right), \tag{16}$$

where M represents a certain function of no importance in this case. $\overline{\omega}$ denotes the product of the angular velocity of the Earth's rotation to the sine of the latitude of given point. But it is easy to see that with $\psi = 0$ the displacement ξ is directed along the parallel and η along the meridian. Therefore, taking the old denominations, we find

$$\frac{\partial u}{\partial t} = \frac{\partial^2\xi}{\partial t^2}; \quad \frac{\partial v}{\partial t} = \frac{\partial^2\eta}{\partial t^2} \tag{17}$$

and consequently

$$\operatorname{tg}\alpha = \frac{\dfrac{\partial^2\xi}{\partial t^2}}{\dfrac{\partial^2\eta}{\partial t^2}}. \tag{18}$$

Having put into (15) and (16) instead of θ its expression (10), performed the differentiation and supposed $\psi = 0$, $r = 0$, we shall find:

$$\operatorname{tg}\alpha = \frac{2\overline{\omega}}{\sigma}. \tag{19}$$

But instead of σ we can introduce its expression in T and instead of $\operatorname{tg}\alpha$ — the value $\frac{n}{m}$, or $\sqrt{\frac{1}{m^2} - 1}$ (what is the same).

Then it will prove that

$$\frac{1}{m^2} = \frac{\overline{\omega}^2\rho^2}{3.66\,c^2} + 1. \tag{20}$$

On the ground of (7) let us exclude the value c^2 from (20). Then, after the transformations we shall get for m the expression

$$m = \frac{2}{q + \sqrt{q^2 + 4}}, \tag{21}$$

where q is the concise form

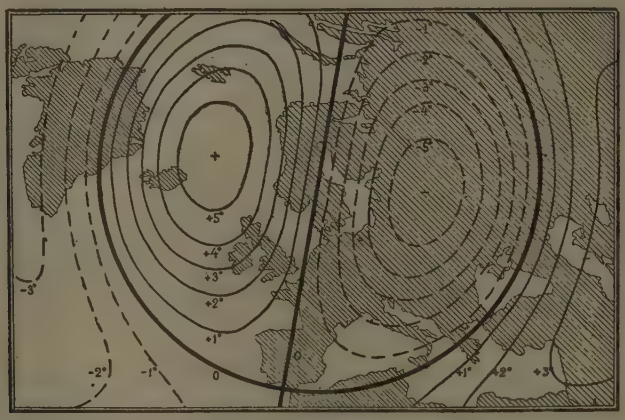
$$q = \frac{\delta \overline{\omega^2} \rho^2}{3.66\, \Pi \Theta}. \qquad (22)$$

Let us place here the probable numeral values $\Pi = 1.6 \cdot 10^3$ [on the ground of the cited work (2)]; $\Theta = 4°$, $\overline{\omega} = 6.3 \cdot 10^{-5}$ (for the latitude 60°), $\rho = 2.5 \cdot 10^8$ (according to considerations given farther in connexion with the schematic chart of isoplates).

Then, according to (21) and (22), it will prove that

$$m = 0.074; \quad \alpha = 85°.7; \quad T \approx 8 \text{ days}.$$

As it was shown above, in the operations with (5) we have supposed that the mean difference of temperatures Θ between the upper (warm) and the lower (cold) zones of the monsoon field is slightly changed with the change of co-ordinates. Actually Θ undoubtedly changes from point to point in the monsoon field. Consequently, the wave velocity c expressed by the formula (7) must be changed too. However, this circumstance



cannot bring any real changes in the conclusions exposed: the latter will only become very complicated with the consideration of the changes of Θ, and our problem will be analogous to that of the seiches in a sea of changing depth. Instead of simple Bessel functions, we shall be obliged to deal with Mathieu compound functions.

For further accuracy of analysis we must renounce to the second assumption, which was introduced in the operations with (5). Namely, we have to consider the fact that a very large portion of heat, determined by the right part of the equation (5), is radiated into the world space and, consequently, does not influence $\frac{\partial \theta}{\partial t}$. But with the consideration of this fact the problem will cease to be lineal. Undoubtedly it will be proved that the oscillations in the monsoon field are maintained according to the same reasons as in other auto-oscillating systems, i. e. at the expense of the current of energy through such a system (in this case at the expense of the transfer of Atlantic heat).

In spite of the simplifying assumptions above-mentioned, our conclusions agree very well with the results of immediate observations: the isoplates, drawn on the schematic chart, remind very closely the outlines

of the isoplates of the divergences of the air temperature from the mean value, drawn by Sandström (4) in conformance to 36 various directions of wind at the Atlantic coast (by the Lofoten).

Examining the material of many years' observations, this author has, to our opinion, put down the different phases of oscillations in the monsoon field, having discarded the influence of various foreign intrusions.

In conclusion we have to note that the localization of nodal lines and loops on the Earth is undoubtedly determined by the distribution of water spaces and surfaces of continents and archipelagoes. In particular, the peculiar «resemblance» between Europe and Greenland becomes quite natural (the tempetature variations at Perm perfectly coincide in phase with those at Upernivik). Equally natural is also the position of the loops of variations in the zone of the Iceland pressure minimum.

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GEOPHYSICS

ON A CASE OF THE PROPAGATION OF RADIO-WAVES IN THE IONOSPHERE

By V. N. KESSENIKH

(Communicated by L. I. Mandelstam, Member of the Academy, 10. I. 1939)

The theory of the magneto-ionic effect in the ionosphere ([1]) leads to an equation determining the possible value of phase velocity of a flat wave. This equation in the absence of absorption can be represented in the following way:

$$\begin{vmatrix} (1-y^2)(1-n^2)-x(1-y^2s_0^2) & -ixyc_0 & xy^2s_0c_0 \\ ixyc_0 & (1-y^2)(1-n^2)-x & -ixys_0 \\ xy^2s_0c_0 & ixys_0\,1-y^2-x(1-y^2c_0^2) & \end{vmatrix} = 0, \quad (1)$$

where $x = \frac{4\pi Ne^2}{m\omega^2} = \frac{\sigma}{f^2}$; N is the number of electrons per 1 cm^3; e and m are charge and mass of electrons respectively; ω and f are circular frequency and number of periods;

$$y = \frac{f_H}{f}; \quad f_H = \frac{eH}{2\pi mc};$$

H is full intensity of the magnetic field of the earth;

$$s_0 = \sin\vartheta, \quad c_0 = \cos\vartheta;$$

ϑ is an angle formed by the direction of propagation and the magnetic field of the earth; $n = \frac{c}{V}$, V is phase velocity.

Equation (1) is of the second power relatively to n^2, while for its roots it has two values of n^2, corresponding to the so-called ordinary and extraordinary rays.

From the viewpoint of ionosphere observations, outstanding interest belongs to the case of propagation in a heterogeneous medium with parallel surfaces of equal ionization in a direction normal to these surfaces. This case corresponds to the usual measurements of critical frequencies at vertical incidence. In fulfilling the condition

$$\frac{1}{V}\frac{\partial V}{\partial z}\lambda \ll 1,$$

it will be possible to apply to this case the calculations referring to propagation in a homogeneous medium.

The condition of reflection with vertical incidence $n=0$ leads to the following equation of the third power relatively to x, which is obtained from (1) if we assume $n=0$:

$$\begin{vmatrix} 1-y^2-x(1-y^2s_0^2) & -ixyc_0 & xy^2s_0c_0 \\ ixyc_0 & 1-x-y^2 & -ixys_0 \\ xy^2c_0s_0 & ixys_0 & 1-y^2-x(1-y^2c_0^2) \end{vmatrix} = 0. \qquad (2)$$

The following three values of x serve as the roots of this equation irrespective of the value of $\frac{s_0}{c_0}$:

$$x_0=1; \quad x_1=1-y; \quad x_2=1+y.$$

In investigating the propagation of the signal it is also necessary, in accordance with Bajpai and Mathur's (2) calculations, to take into account an additional fourth root corresponding to the transformation of group velocity U into zero.

This condition reads as follows:

$$\frac{d}{d\omega}(\omega n)=n-\omega\frac{dn}{d\omega}=0. \qquad (3)$$

The fourth root of the equation is, generally speaking, close to x_0.

As regards the question, whether the reflected components of the signal correspond to all the enumerated values of x, there still exists great variety of opinions.

Kirby, Berkner and Stuart (3), as well as Berkner and Wells (4), claim that the reflection corresponding to the root of $x_2=1+y$ cannot take place, since the ray in the region of the given concentration always breaks up into only two rays — an ordinary one and an extraordinary one. The extraordinary ray is fully reflected at a height corresponding to the least concentration of $x=1-y$; only the component with a polarization corresponding to the ordinary ray is apt to penetrate higher. This component, in the opinion of the authors mentioned, is fully reflected when $x=1$.

Other authors, such as Taylor (5), Försterling and Lassen (6), point to the possibility of a reflection corresponding to a third root too: $x_2=1+y$.

Without dwelling upon this question at any considerable length, we may bring forward the fact that the splitting up of the signal into three components and the three respective values of critical frequency have been systematically observed by a number of investigators of the ionosphere [cf. for instance the article by Darshi Bansal (7)].

According to the data obtained by the Tomsk ionosphere station, three components are rather regularly detected, particularly in winter at night.

The course of reasoning followed by Berkner and Wells and others is erroneous, as they fail to account for the change of the multiplier of polarization for the ordinary and extraordinary ray with change in x. With $x=1$, the ray, which for $x=1-y$ appeared by its polarization to be an ordinary one, is now found to consist again of two components, of which one is reflected when $x=1$, while the second one passes up to the level $x=1+y$, where it is partly reflected.

To the three reflected components correspond three critical frequencies which, on the assumption of the constancy of H along the path of propagation, are to be determined according to (3) as

$$\left.\begin{aligned} f_0 &= \sqrt{\sigma_{\max}}, \\ f_1 = f^x &= \frac{f_H + \sqrt{f_H^2 + 4\sigma_{\max}}}{2}, \\ f_2 = f^z &= -\frac{f_H + \sqrt{f_H^2 + 4\sigma_{\max}}}{2}. \end{aligned}\right\} \qquad (4)$$

The difference of the two values of critical frequency of the extraordinary ray should be

$$f^x - f^z = f_H. \qquad (5)$$

For Tomsk $f_H = 1.5$ mc/sec.

The observational data of the Tomsk ionosphere station show the difference $f^x - f^z$ to increase in a number of cases up to 2.2 mc/sec. and to fall down in many instances to 1 mc/sec., fluctuating on the average about the values of 1.5 mc/sec.

These fluctuations cannot be explained on the basis of equation (5). $f_H = \frac{He}{2\pi mc}$ may change only with change of H. The relative change of H during the most severe magnetic storms is never liable to exceed 5%.

The fluctuations of $f^x - f^z$, observed in reality, have a considerably greater magnitude. Among the possible reasons for the change of $f^x - f^z$, the supposition may be framed regarding the influence of the non-homogeneity of the magnetic field of the earth along the path of the ray in a refracting region.

We shall examine the case of propagation within a heterogeneous magnetic field, when the relatively small heterogeneity of the magnetic field of the earth may involve a considerable increase of the difference of critical frequencies of the extraordinary ray, an increase many times exceeding the relative change of f_H.

Horizontal currents in the conducting strata of the atmosphere produce variations in the terrestrial magnetic field. In order to elucidate the qualitative aspect of the effect we are concerned with, we shall confine ourselves to a consideration of the simplest scheme of the distribution of ionization as well as the additional current field according to height.

Let us assume that the additional current field is symmetrically distributed with reference to the level of maximum ionization. Counting off the height, z, from the level of maximum ionization, we may represent dependence σ upon the height in the following way:

$$\sigma = \sigma_0 - \beta z^2 \qquad (6)$$

and f_H as

$$f_H = f_{H0} + \alpha z, \qquad (7)$$

which would hold in the case of a horizontal direction of the basic field (more complex cases may be examined in a similar way).

Substituting these expressions for σ and f_H into equation (3), we obtain the following solutions for z, corresponding to the height of the reflection of an ordinary ray and two components of an extraordinary ray (discard-

ing the positive values of z, as corresponding to the negative gradient of ionization):

$$\left.\begin{aligned} z_0 &= -\frac{1}{\sqrt{\beta}}\sqrt{\sigma_0 - f^2}, \\ z_1 &= \frac{\alpha f - \sqrt{\alpha^2 f^2 + 4\beta(\sigma_0 - f^2 + f f_{H0})}}{2\beta}, \\ z_2 &= \frac{-\alpha f - \sqrt{\alpha^2 f^2 + 4\beta(\sigma_0 - f^2 - f f_{H0})}}{2\beta}. \end{aligned}\right\} \quad (8)$$

From the condition of the materiality of z, we obtain three values of the critical frequency:

$$\left.\begin{aligned} f^0 &= \sqrt{\sigma_0}, \\ f^x &= \frac{f_{H0} + \sqrt{f_{H0}^2 + 4\sigma_0\left(1 - \frac{\alpha^2}{4\beta}\right)}}{2\left(1 - \frac{\alpha^2}{4\beta}\right)}, \\ f^z &= \frac{-f_{H0} + \sqrt{f_{H0}^2 + 4\sigma_0\left(1 - \frac{\alpha^2}{4\beta}\right)}}{2\left(1 - \frac{\alpha^2}{4\beta}\right)}, \end{aligned}\right\} \quad (9)$$

whence

$$f^x - f^z = \frac{f_{H0}}{1 - \frac{\alpha^2}{4\beta}}. \quad (10)$$

During severe magnetic storms α can reach magnitudes as high as $10^4 \frac{1}{\text{kmsec.}}$; β is usually ot the order of 10^{10}, but in the process of effacing the maximum of ionization it is liable to decline considerably, in consequence of which cases are possible when $\frac{\alpha^2}{4\beta}$ will approach unity. In these cases an appreciable increase of $f^x - f^z$, as compared with f_H, ought to take place.

Thus, with strong effacing of the ionized region, which is in particular observable during magnetic storms, there exists the possibility of an increase in the difference of critical frequencies of the extraordinary ray at the expense of the relatively small heterogeneity of the magnetic field of the earth.

From the example examined it follows that a complete interpretation of the given ionosphere observations necessitates an evaluation of the influence of various combinations of the gradient of ionization and the heterogeneity of the magnetic field of the earth.

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BIOPHYSICS

COMPARATIVE DYNAMICS OF SEA ANIMALS

By W. W. SHOULEJKIN, Corresponding Member of the Academy,
V. S. LOUKJANOVA and I. I. STASS

In one of the preceding works [1] the authors have investigated the hydrodynamic properties of some fish and the dolphin in not too wide limits of the variations of Reynolds' number. In the same work there was suggested that apparently the specific power of sea animals (i. e. maximum power referred to 1 g of the body mass) may be taken as constant. Proceeding from the same suggestion, we had to assume that the maximum velocity attainable for a fish has to grow proportionally to the root of a certain power (approaching the third one) of the lineal dimensions of the body.

This assumption was grounded on the condition of rather slight variations of Reynolds' number. In the present work the limits of the investigations were extended in such a form that it proved possible to connect the elements of the dynamics of fish (adult ones and fry) with those of whales and of plankton copepode crustaceans.

Primarily we have made an investigation of the changes of the dimensionless resistance coefficient: $\psi = \frac{2f}{\delta \cdot S \cdot v^2}$ under the variations of Reynolds' number $R = \frac{d \cdot v}{\nu}$ in very wide limits.

It is known that in reduced formulae f denotes the force resisting the movement of a solid in the liquid with the density δ with the velocity v. S represents the maximum area of the cross-section, d, the lineal transverse dimension of the body in this section and ν, the kinematic viscosity of the liquid.

The fish were taken of the most streamlined shape [contrary to the work cited [1]]. The measurements corresponding to particularly large Reynolds' numbers (exceeding 10^5) were made on a specially built wooden model of a fish, which allowed to change the leaden weights without changing the body outlines. For particularly small values of Reynolds' number having both theoretical and practical interest (about 10^{-1}), metal models (aluminium, brass and leaden ones) falling into castor oil were used. For better control over the accuracy of manufacturing, these models had a perfectly circular cross-section (and not an elliptic one which most fish have). Therefore on the basis of the curve segment got for small models (denoted by № 5 in the diagram of Fig. 1) we had to draw the respective segment of the characteristic curve sought for by means of graphic extrapolation.

The characteristic curve sought for, plotted from direct measurements of fish, of the large model and from the above extrapolation is shown in Fig. 1 under № *1* (the solid curve). For comparison, in this figure there have been given four more characteristic curves, corresponding to the motion of a sphere (*2*), of a round cylinder, for which the resistance power was referred to the length equal to the diameter (*3*), of a cylinder with the section of streamlined shape conformably to the same length (*7*) and of an ideally streamlined solid, the length of which three times exceeded the diameter of the largest section (*6*). These four curves were taken from a reference book on physics ([2]), where they had been plotted from a series of experiments.

We can see that the advantages of the streamlined shape of the fish body, as compared with the spheric one, begin to appear only with the value of Reynolds' number exceeding 10^4. This may be the cause responsible for slow-moving fish swimming among sea-weeds, having assumed in the process of evolution a shape reminding that of a sphere.

Above Curve *1*, corresponding to fish and to whales of fish-like shape, we see Curve *4*, representing the resistance coefficient of Copepoda crustaceans (*Paracalanus*, *Centropages*). This curve was plotted on the basis of experiment which will be described in a special article (by the second author of the present work). The experiments were performed with models of bodies made on a highly increased scale, and therefore in order to realize the law of similarity the normal Reynolds' number was attained by means of replacing sea water by a higher viscous concentrated sugar solution.

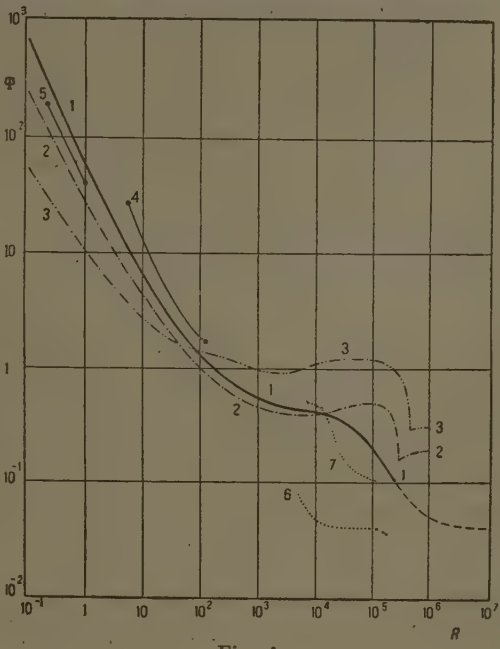

Fig. 1.

After the characteristic curves for fish-like bodies (*1*) and for plankton crustaceans (*4*) had been obtained, it proved possible to trace the comparative dynamics of sea animals much farther than it had seemed at the beginning of our investigation.

It is easy to see that the energy necessary to overcome the force f of water resistance will be expressed

$$W_v = \frac{1}{2}\psi \cdot \delta \cdot S \cdot v^3. \qquad (1)$$

Suppose that the efficiency of the propeller of the fish is equal to η, that its body mass M is expressed through the area S of the maximum cross-section, the length of the body L and a certain mean density of the body δ, practically approaching the density of water:

$$M = \beta_1 \cdot \delta \cdot S \cdot L. \qquad (2)$$

Then the energy spent by muscles on the surmounting of the external forces referred to 1 g of the body mass will be

$$w_v = \frac{\psi}{2} \cdot \frac{v^3}{\beta_1 \cdot \eta \cdot L} \cdot \qquad (3)$$

In addition to this, a certain amount of energy must be spent to overcome internal friction in the muscular substance itself. If we denote the relative deformation of a muscle, i. e. the change of one linear cm of muscular fibril, through σ, the work of friction inside 1 g of muscle under elementary deformation $d\sigma$ will be, broadly speaking, expressed in

$$dP = \frac{\mu_m}{\delta_m} \cdot \frac{\partial \sigma}{\partial t} \cdot d\sigma. \qquad (4)$$

Here μ_m represents the coefficient of friction of the muscle, δ_m, its density. But the law of the oscillations of σ referred to time had been found for the fish by the first author of the present article ([3]). Having applied it in this case and performed some simple transformations, we shall get the following expression for the power absorbed by the tenacity of muscles, referred to 1 g of the body mass:

$$w_r = \frac{2\pi^2}{T^2} \cdot \nu_p \cdot k^2. \qquad (5)$$

Here T denotes the period of transverse oscillations of the body; ν_1, the value having the character of the kinematic viscosity of the muscles, and the expression k is: $k_0 = \frac{\overline{A \cdot b}}{\lambda^2}(4\pi^2 + \Delta^2)$, where in its turn $\overline{Ab}$ represents the mean product of the amplitude of the transverse oscillations of the body to the distance of the muscular fibril from the axis of the spine; λ denotes the length of the «solid wave» passing along the body of the fish during the movements; Δ, the logarithmic increment characterizing the growth of the amplitudes of the transverse oscillations from head to tail. In the same paper ([3]) it had been pointed out that the ratio of the way vT, passed by the fish in one period of oscillations, to the length of the body L is a certain characteristic value which may be named the «reduced step». Let us express it shortly

$$\frac{vT}{L} = \beta_3. \qquad (6)$$

Then on the ground of (5) and (6) we can write down:

$$w_r = \frac{2\pi^2}{\beta_3^2} \cdot k^2 \cdot \nu_p \cdot \frac{v^2}{L^2}. \qquad (7)$$

All the energy spent by a fish on the surmounting of external and internal resistances referred to 1 g of the mass will be expressed on the basis of (3) and (7) in the form

$$w = \frac{\psi}{2} \frac{v^3}{\beta_1 \cdot \eta \cdot L} + \frac{2\pi^2}{\beta_3^2} \cdot k^2 \cdot \nu_p \cdot \frac{v^2}{L^2}. \qquad (8)$$

Let us call it the specific power of the animal.

Out of (8) we can deduce the expression of the velocity v of the movement of the fish from the length of its body L:

$$v = \frac{n \cdot L^{1/3}}{\left(\psi + \frac{\varkappa}{R}\right)^{1/3}}, \qquad (9)$$

where

$$n = (2\beta_1 \cdot \eta \cdot w)^{1/3}, \qquad (10)$$

$$\varkappa = 4\pi^2 \cdot \frac{\beta_1 \cdot \beta_2}{\beta_3^2} \cdot k^2 \cdot \frac{\nu_1}{\nu}. \qquad (11)$$

In its turn ν denotes the kinematic viscosity of water and R, as previously, the Reynolds' number; $\beta_2 = \frac{d}{L}$, where d is the diameter of the maximum cross-section of the body.

The formulae (9), (10) and (11) will become particularly important in the case, if we understand w as the maximum specific power attainable for the animal. Then, the formula (9) would, obviously, express the maximum velocity possible for an animal of the given length L.

Proceeding from physical and chemical considerations, we can expect to find that the maximum specific power changes very slightly from fry to a gigantic whale. Let us point out that all the other values included in the expressions (10) and (11) change apparently in very close limits with the change of the body dimensions. Therefore the values of their third roots will change very slightly. Lacking at the present time detailed investigation, we shall consider n and $\varkappa$ constant and extend the formula (9) to the fast-moving whales, the motion of which is also operated by means of «solid waves» propagated along their body.

Let us try to prove the validity of such assumptions by comparing the results of direct measurements (for different sea animals) with the theoretic formula (9). As the latter contains two constants, the corresponding curve is determined by two given points. However, the expression (9) in itself is inconvenient for the drawing of curve. Therefore we shall get out of it two equations, which will allow us to determine the connexion between v and L in a very simple parametric form.

Let us exclude L from (9), expressing this value by means of R and v:

$$L = \frac{\nu \cdot R}{\beta \cdot v}. \qquad (12)$$

Then, after some transformations, we shall get instead of (9):

$$v = U \left(\frac{R}{\psi + \frac{\varkappa}{R}} \right)^{\frac{1}{4}} \qquad (13)$$

and

$$\frac{\beta \cdot U^4}{\nu} = n^3. \qquad (14)$$

The constants U and $\varkappa$ will be determined, setting before us the numbers characterizing the movement of the grey mullet fry, on the one hand, and of the seventeen-meter-long whale (*Balaenoptera borealis*), on the other.

The measurements made at the Black Sea Hydrophysical Station have given for the first object: $L = 1.4$ cm; $v = 40$ cm/sec.; $R = 1.1 \cdot 10^3$. For the seconde one, according to literary data: $L = 1700$ cm; $v = 1500$ cm/sec.; $R = 4.5 \cdot 10^7$.

Having calculated the constants U and $\varkappa$ and using the diagram of Fig. 1 (Curve *I*), it is easy to determine according to (13) the maximum velocities v corresponding to different values of the parameter R. After this, on the ground of the formula (12), the values of L corresponding to these maximum velocities will be determined.

The theoretical curve drawn in this way (according to two empiric points) is shown in the diagram of Fig. 2, where the top limit is the

point corresponding to the whale *Balaenoptera borealis*. In the direction of small lengths L the curve is extrapolated up to the values of L corresponding to the dimensions of the smallest fry existing.

We can see that the curve fits quite satisfactorily the points obtained by means of direct measurements by the third author of this article (made with fish swimming in a large basin). Only one point, correspond-

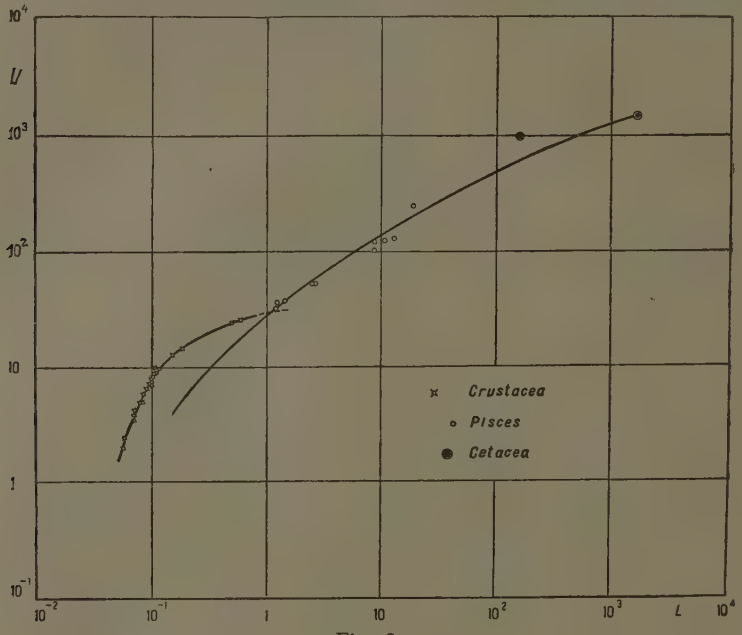

Fig. 2.

ing to the dolphin [velocity, according to Gray (4)], lies considerably above the curve, the fact finding its explanation in certain extraordinary properties of the dolphin energetics. The chief hypothesis of the constance of w can, consequently, be considered as corroborated.

In the diagram of Fig. 2 another shorter curve is given, passing across the points obtained by means of direct observations of plankton copepoda crustaceans: *Paracalanus* and *Centropages* (the dominating number of small circles in the diagram) and of several crustaceans of the family *Gammaridae* (the four points in the right part of the short curve).

We see that below a certain limit value of L the fish begin to lag behind the Copepoda in maximum velocities of movement. That is why in the process of evolution the fish appeared only at a certain stage of the increase of length. This stage was preceded by a period invaded by the Copepoda. Still earlier, the ciliates seem to have been the most rational propeller.

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GEOCHEMISTRY

INDIUM IN THE WASTE OF THE TCHIMKENT LEAD WORKS

By I. P. NOVOKHATSKY and S. K. KALININ

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In the paper by S. A. Borovick and S. K. Kalinin entitled «A Spectroscopic Investigation of the Composition of the Products of Smelting of Lead and Zinc Plants» (1) mention was already made of the increased contents of indium in certain products of the Tchimkent lead works.

In the present paper we shall endeavour to discuss in detail the distribution of indium by separate products in connexion with the separate stages of the metallurgical process.

The Tchimkent lead works at present receive the main bulk of their raw material supply from the ores of Karatau and Kara-Mazara. The presence of indium has been found only in Kara-Mazara ores, the amount of this element reaching here up to 0.001%. In these ores indium is genetically bound up with zinciferous minerals, part of which reach the works together with lead-bearing material. No indium has been found in the concentrates of Karatau (mines Atchi-Say and Kantagi).

In the process of the pyrometallurgical treatment of the ores, indium has been found to behave with considerable complexity. At the same time, concentration of indium is observed to take place at definite stages of the process.

In the first stage of the process—in sintering—indium partly escapes with the gases in the form of dust which later settles down in the Cottrell electrostatic precipitation apparatus. This loss is, however, inconsiderable. It would seem that the temperature of sintering (about 700—750°) is insufficient for volatilization. The indium content in Cottrell dust of the sintering section approaches to that in ores.

In course of the next process—smelting for work-lead which is effected at higher temperatures (about 1100—1200°)—indium again partly volatilizes and is deposited with dust in the Cottrell apparatus. Another part, in course of smelting, becomes dissolved in work-lead and is almost completely separated from the lead in course of the process of igneous refining, being removed with lead and rich slickers. Thus, appreciable concentration of indium are observed in the Cottrell dust of big blast-furnaces and in slickers (dry and fat).

In the other products—matte and disposable slag (still containing metal values)—indium is observed in the spectrographs only in the form of very faint lines. No more indium is to be observed in the further products of smelting. The slickers enriched with indium are remelted for work-lead in a small blast-furnace provided with a special Cottrell dust-catcher. Investigations

of the dust of the small blast-furnaces have shown a maximum content of indium, as compared with the other products, reaching in separate samples as high as 0.1%.

It is necessary to note that the content of indium in the various products of the metallurgical treatment is far from constant, being subject to considerable variations depending on the character of the charge. In the working of ores from the Karatau region alone, no indium is to be found, while in those cases when the plant is working only with Kara-Mazara ores, there will be observed in the products of the treatment a definite increase in the amount of that element.

Within the very next years a number of new deposits will begin supplying the Tchimkent works with ore: Tekeli (Djungar Alatau), Ack-Tjuss and Bu-Urdu (Northern Kirghizia). The latter two deposits appear to be of particular interest with respect to indium content; separate samples from these deposits contain up to 0.1% indium.

It stands to reason that the content of indium in the products of smelting of the Tchimkent works will also increase as soon as these ores begin to be used.

In connexion with the practical feasibility of obtaining indium from the waste of the Tchimkent works, the question ought to be raised already now of smelting separately Kara-Mazara and Northern Kirghizia ores, on the one hand, and Karatau ores, on the other hand, this being all the more necessary, as the mixing of ores in the charge is apt to bring in a number of objectional impurities (such as As, Sb, Bi), which are found to be very resistant in the process of smelting, penetrating even into the final product and reducing its quality.

In the further smelting of slickers in the small blast-furnace, where a number of additions containing no indium are made, the resultant dust becomes somewhat leaner in respect of indium content. Taking this into consideration, it might be advisable to effect the treatment of slickers separately from all the other products in a special furnace equipped with a dust-catching apparatus. Speaking on the whole, we find that the stated content of indium in the dust of the small blast-furnace is in comparison both with the other products of the Tchimkent works and the percentage share of the products of other plants altogether exceptional and deserves the most careful attention.

Further research work should be directed towards studying the conditions of the concentration of indium in dust and also the methods of its extraction.

As to the question why it is notably in the slickers that the concentration of indium is so considerable, only tentative answers can so far be given to it. In the processes of manufacturing operations, indium behaves much in the same way as when it is confronted with natural conditions, i. e. this element becomes concentrated at high temperatures and also displays a definite relationship with zinc, this latter property being well known to us from our knowledge of zinc-blende.

Slickers contain more zinc than other products.

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BIOCHEMISTRY

TREND OF INVERTASE ACTIVITY IN FRUIT OF CUCURBITS

By V. V. ARASSIMOVICH

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The synthesis of organic substances in the cell (for instance, of sucrose in leaf root or of starch in potato tubers) is usually supposed to go on under the influence of respective enzymes. But not until recent years and until the method of vacuum infiltration was introduced into practice could the study of the synthetic activity of enzymes in a living cell attain its present extensive development. This method was described by Björksten ([2]), applied to the investigation of proteolysis by Mothes ([6]) and, after being elaborated by Kursanov ([4]), was used widely in this country for the determination of synthetic and hydrolytic activity of the invertase. It turned thereby out that the relation between these two activities of the invertase goes parallel with the presence of a number of physiological characters in plants. Many well-founded data in support of this connexion have been adduced by A. I. Oparin ([7]), A. L. Kursanov ([4], [5]), B. A. Rubin ([8], [9], [10]), N. M. Sisakian ([11]) for some vegetable species and varieties.

The present work has employed the method of vacuum infiltration in order to reveal the differences as regards invertase between genus *Citrullus* and genus *Cucumis*, as well as between their species and varieties. The work was carried on in 1938 near Tashkent. The material for the experiment was supplied by *Cucurbitae* section of All-Union Institute of Plant Industry.

The species and varieties of melon are highly variable both in total and relative content of sugars in their fruit. Along with species entirely destitute of sucrose there are such as are higher in it than in other sugars.

The action of the invertase was being studied by us immediately on fruit flesh-sections. Mention needs to be made of considerable difficulties experienced in applying the method of vacuum infiltration to cucurbital fruit, especially that of table varieties of water-melon, which is high in water and of loose anatomical structure. On that account the experimental data can scarcely be credited to present absolute values. In the slices, about 1.5—2 mm thick, 0.1 mol solution of sucrose and 0.2 mol solution of monosaccharides were infiltrated. On the average 100—250 mg and in some cases 400—500 mg of sugars were introduced in that way into each of the 10-g samples. The value used throughout this work to express the synthetic and hydrolytic activity of the invertase is the percentage of sugars inverted during 3 hours out of the total of sugars introduced by filtration.

The species of water-melon can be divided properly into two groups according to their total and relative content of sugars. The fruits of the first group

are high in sugars and contain sucrose. Such are the widespread table water-melon *C. edulis* and the rarely occurring *C. fistulosus*. The predominance of the fructose over the glucose is a characteristic feature of the table water-melon, while there is either equal proportion of both in *C. fistulosus* or else it is the glucose which predominates.

The second group of water-melons includes the species *C. colocynthoides* (fodder water-melon) and *C. colocynthis* (wild African water-melon). These are extremely low in sugars (not above 1.5%, which ratio drops to 0.3—0.4% in *C. colocynthis*) and contain no sucrose. The glucose predominates over the fructose in both species.

Applying the vacuum infiltration of monosaccharides and sucrose, we did not fail to detect a synthetic and hydrolytic activity of the invertase in ripe fruits of *C. edulis* and *C. fistulosus*, containing sucrose. The only conclusion to be drawn from the experiments referred to in Table 1 is that both the synthetic and the hydrolytic invertase are active in the two species tested.

Different picture was observed in *C. colocynthoides* and *C. colocynthis*. In ripe fruits of these species, which contain no sucrose and are low in sugars, no synthetic action of invertase could be detected, while the hydrolytic invertase was only active as the whole of the sucrose introduced was inverted to monosaccharides.

The results of our analyses are summarized in Table 1.

Table 1

Ratio of Sugars and Activity of Invertase in Various Water-Melons*

Water-melons	Sugar ratio in % of raw weight			Activity of invertase (ratio of inverted sugar to introduced sugar, in %)	
	total	monosaccharides	sucrose	synthesis	hydrolysis
C. edulis	7.98	3.64	4.34	9	83
	10.00	4.25	5.74	17	0
	5.34	4.25	1.09	38	14.6
	6.60	4.87	1.72	65	7
C. fistulosus No 1826	6.31	4.51	1.80	48.8	86
C. colocynthoides No 197	1.10	1.10	0	0	100
	1.32	1.32	0	0	100
C. colocynthis No 420	0.50	0.50	0	0	50
	0.33	0.33	0	0	100
	0.35	0.35	0	0	50

Naturally enough, the activity of invertase, as it is also apparent from the data obtained, is correlated with the presence or absence of sucrose in the species tested as well as with their relative content of monosaccharides. Table varieties of water-melon are high in fructose which is probably form-

* The data in this and in the following table are obtained on single fruits and it is only on accout of the smallness of fruit in *C. fistulosus* and *C. colocynthis* that the analyses of these species have been made on samples taken from 2 or 3 fruits.

ing on the continuous hydrolysis of sucrose. On the other hand, the fodder and the wild varieties which are lacking sucrose are consequently low in fructose, since the latter cannot possibly accumulate on hydrolysis. And that it is still present in small amounts should be ascribed to the conversion of the glucose into fructose which process is continuously going on in the cell.

Comparative estimation of the invertase action in ripe and unripe fruits has brought into evidence that in *C. edulis* and *C. fistulosus* the activity of the invertase proceeds in both directions whatever the ripeness of the fruit. Nor is it definitely related to ripeness in the fodder and in the wild varieties, in neither of which could we detect any synthetic activity while the hydrolytic activity was very intense in both.

The sucrose introduced into the fruits was hydrolyzed entirely. It is worthy of note that the same was observed in the insipid form of *C. colocynthis*. Only in one variety of the fodder melon did we reveal a weak synthetic activity of the invertase in unripe fruits, along with its high hydrolytic activity. No ripe fruits of that variety were available for the test.

We may thus state that in *C. edulis* and *C. fistulosus*, which contain sucrose, invertase is active both in synthesis and hydrolysis, while in *C. colocynthoides* and *C. colocynthis* the hydrolytic activity of the invertase is sharply expressed and the synthetic is either entirely absent or so weak as not to be detected by the method of infiltration.

Interspecific hybrids of the water-melon, as we have formerly stated (1), show unmistakable tendencies to approach the less cultivated initial form as regards total and relative content of sugars. The same results were obtained when the data on invertase activity in the fruits of F_1 from interspecific crosses were compared with those for the parent forms.

Table 2

Inheritance of Sugar Content and Invertase in Interspecific Crosses between Water-Melons

Parent forms and hybrids	Sugar content in %			Activity of invertase (ratio of inverted sugar to introduced sugar, in %)	
	total	monosaccharides	sucrose	synthesis	hydrolysis
C. edulis	5.34	4.25	1.09	38	14.6
	6.60	4.87	1.72	65	7
C. colocynthis	0.50	0.50	0	0	50
	0.33	0.33	0	0	50
				0	100
F_1 *C. edulis* × *C. colocynthis*	2.50	2.50	0	0	50
	—	—	—	0	100
C. edulis (another variety)	7.98	3.64	4.34	9	83
	10.0	4.25	5.74	17	0
C. colocynthoides N 197	1.10	1.10	0	0	100
	1.32	1.32	0	0	100
F_1 *C. edulis* × *C. colocynthoides*	3.26	3.02	0.23	26	100
	—	—	—	0	86

As seen from the tabulated data, the invertase of table varieties is active both synthetically and hydrolytically. But the hydrolytic action predominates

in F_1 hybrids obtained by crossing these varieties with the wild and the fodder water-melons, where only the hydrolytical activity of the invertase comes into play. The differences with regard to invertase between the parent forms and the F_1 hybrids go parallel with the presence or absence of sucrose in them.

In unripe fruit of the F_1 hybrid *C. edulis*×*C. colocynthis* there was no synthetic activity of the invertase either, while its hydrolytic activity was very high. We see thus that the ratio of synthetic to hydrolytic activity proper to the invertase of *C. colocynthis* is predominant in the fruit of F_1 hybrid, ranging in condition between small ovary and full ripeness.

The invertase of the musk-melon was studied on species *C. eu-melo* Pang., to which all table varieties of musk-melon refer, and on *C. chinensis*, which yields small unsweet fruit. As in the case of water-melon, it has been established that the table varieties of musk-melon, which are high in sugars and, especially, in sucrose, show a predominance of the synthetic activity of the invertase over its hydrolytic activity, the latter being very low or entirely absent. In unsweet musk-melons *C. chinensis*, in which no sucrose is available, only hydrolytic activity could be detected.

Among the varieties of the species *C. eu-melo* some definite differences in the character of the invertase have been recorded between good and bad keepers (winter melons and summer melons).

From comparing our observations on the activity of the invertase in the musk-melons and water-melons we can explain with a certain degree of precision the conception advanced by N. N. Ivanov and R. S. Alexandrova([3]) as early as in 1929 about the dynamics of sugars in cucurbital plants on their ripening.

Part of the glucose conveyed from the leaves to the fruit is converted into fructose, and in green fruit of cucurbits both monosaccharides are always present, but the former of them predominates. The sucrose is built up in the fruit by means of the invertase. At the outset, when both synthesis and hydrolysis are active with the preponderance of the latter, there is but very little sucrose or it is altogether undetectable. But with the time the hydrolysis gradually weakens; in musk-melons it vanishes entirely and sucrose becomes predominant over other sugars in the ripe fruit.

The results of the experiments on cucurbital plants corroborate the observations made by A. L. Kursanov and B. A. Rubin on other plants, viz., that the action of invertase is definitely related to the presence or absence of sucrose. The most representative example of this relationship is afforded by the case of interspecific hybrids of the water-melon. But this dependence holds true only when different species are compared, and within the same species of cucurbital plants no such definite correlation between the value of synthesis and the amount of sucrose is observable.

A similarity with regard to the action of the invertase has been detected between winter musk-melons (good keepers) and the fodder water-melons (which keep during 1—2 years). In the latter enhanced hydrolytic activity of the invertase and the absence of synthetic activity was stated before this water-melon was deposited for preservation. In good keeping musk-melons the fermentative activity of the invertase is reduced and the hydrolysis alone is active.

No data are available on the dynamics of the invertase in the winter musk-melons and fodder water-melons after their storage. From the fact that those musk-melons increase in sugars (hydrolysis of polysaccharides) and grow high in sucrose, we may suppose that the synthetic activity of the invertase goes up on storing, along with the hydrolysis of other carbohydrates which proceeds under the action of other enzymes.

The fodder water-melon decreases in sugars on storing and forms no sucrose. Consequently, we cannot expect any synthetic activity of the invertase during its storage.

Conclusions

1. By infiltrating sugars into fruit slices of water-melons we have been successful in establishing that the presence of sucrose in table variety and its absence in the fodder variety tested goes parallel with the presence or absence of synthetic activity of the invertase.

2. F_1 hybrids obtained by crossing *C. edulis* with *C. colocynthis* showed neither any presence of sucrose nor any synthetic activity of the invertase.

3. The same has proved to be true of musk-melons. The unsweet form *C. chinensis*, deprived of sucrose, showed only the hydrolytic activity of the invertase, while in *C. eu-melo*, which is high in sucrose, the invertase was almost exclusively active in synthesis.

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MINERALOGY

CLASSIFICATION OF THE MINERALS OF THE SERPENTINE GROUP

By N. E. EFREMOV

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Since the appearance of A. E. Fersman's ([1–2]) monograph on investigations in the domain of magnesian silicates, there have been no substantial works on this subject. At the same time magnesian silicates and particularly the serpentine group, which is one of the least studied ones, should draw more attention in this country, since their study is connected with problems of a scientific practical nature ([4]).

At present we can make note of a definite series of serpentine minerals: hydrophorsterite ([4])—$2MgO \cdot SiO_2 \cdot nH_2O$, serpentine—$3MgO \cdot 2SiO_2 \cdot 2H_2O$, karachaite—$MgO \cdot SiO_2 \cdot H_2O$ ([5]). In this connexion it is of interest to note a new mineral species of the serpentine group, the existence of which in nature was to be expected from our former conceptions on the composition of the serpentine group ([6]).

At the magnesian silicate conference Afanassiev reported that, while studying the olivinites in the region of Saig Lake on the Kola peninsula, he quite often observed veins of a white mineral, sometimes with a slight pale yellow tint. We thought it interesting to study this mineral which we got through the kind help of engineer Afanassiev. In lumps the mineral is white in colour, generally having a pale yellow or cream tint. There are compact and fibrous varieties.

Specific gravity 2.401. Hardness between 2 and 3. Under the microscope the mineral is greyish-white with a cream tint and presents an aggregate of variously orientated scales, laminae and tablets, forming a background, against which comparatively larger individuals may be observed. The latter resemble chrysotile cords. Straight extinction is observed throughout with regard to the cleavage cracks which stand out clearly in the manifold forms, in which this mineral is apt to appear. The mineral is biaxial, positive. The interferential colour is usually light-grey, and seldom rises to yellowish tints of the first order$\sim$0.01; N_m=1.542. The data obtained from an analysis of the compact variety of the mineral, made by T. M. Jakovleva, are given in the table under № 1.

The results of the analysis give a comparatively simple formula for the mineral, which is assumed to be $5MgO.4SiO_2 \cdot 4H_2O$ with $MgO:SiO_2 = 5:4 = 1.25$. A thermal analysis of the mineral, performed after the differential method with the Saladin Le Chatelier apparatus (thermogram № 1605), showed on the whole the endothermic and exothermic effects that are typical

Oxides	№ 1		№ 2	
	Weight per cent	Molecular ratio	Weight per cent	Molecular ratio
SiO_2	43.78	3.95	44.13	2.00
Al_2O_3	0.56	0.03	0.29	—
Fe_2O_3	1.17	0.04	1.73	0.03
FeO	none	—	—	—
CaO	0.41	0.05	—	—
MnO	traces	—	0.23	0.01
MgO	37.01	5.00	38.91	2.62
NiO	0.11	—	—	—
CO_2	0.56	0.07	—	—
H_2O+	13.04	3.94	14.42	2.13
H_2O-	3.02	0.82	1.00	0.15
	99.66	—	100.71	—

of serpentine minerals. A powdergram of the mineral was taken in the X-ray laboratory of Lomonossov Institute of the Academy of Sciences. For comparison sake we took powdergrams of the silky chrysotile-asbestos from the Markopidge region, deveilite from the Tkhacha region and kerolite from the Beden region. The result* was that for all of the four minerals the presence of typical «serpentine» lines was established. Therefore with regard to its physico-chemical properties the mineral in question is a new representative of the serpentine mineral group, and we have called it kolskite after the region where it was discovered. The value of the silico-magnesial module, 1.25, determines its place paradbetween eveilite and karachaite in the table of serpentine minerals given below. It is noteworthy that the mineral, which represents the Caucasian fibrous variety of serpentine (see analysis 2 in the table) and which has already been described by us ([6]), has the same formula $5MgO \cdot 4SiO_2 \cdot 4H_2O$.

We can thus ascribe the following composition to the members of this series, which are known up to the present date:

		MgO : SiO_2
1. Hydrophorsterite**	$2MgO \cdot SiO_2 \cdot nH_2O$	2.00
2. Chrysotile, short-fibred from the Urushten region, North Caucasus([7])	$7MgO \cdot 4SiO_2 \cdot 5H_2O$	1.75
3. Adigeite from the mount Tkhach region, North Caucasus ([6])	$5MgO \cdot 3SiO_2 \cdot 3.5—4H_2O$	1.66
4. Serpentine (theor. formula)	$3MgO \cdot 2SiO_2 \cdot 2H_2O$	1.50
5. Deveilite—$H_{12}Mg_4Si_3O_{16}$ and paradeveilite	$4MgO \cdot 3SiO_2 \cdot 3H_2O$	1.33
6. Ishkildite (chrysotile) from Ishkildino, Urals([8])	$5.25MgO \cdot 4SiO_2 \cdot 3.5H_2O$	1.30
7. Kolskite from the olivinites of the Kola peninsula	$5MgO \cdot 4SiO_2 \cdot 4H_2O$	1.25
8. Karachaite, Shaman-Beklegen	$MgO \cdot SiO_2 \cdot H_2O$	1.00
9. Kerolite (theor. formula)	$MgO \cdot SiO_2 \cdot nH_2O$	1.00

It is important to note that the kerolite minerals also fit in with the general regularity of this series***.

* The results of measuring the powdergram are not given here on account of the space available.

** In the assumed formulae we give approximate oxide coefficients.

*** According to our investigations, the composition of kerolites corresponds to the formula $3MgO.m3SiO_2.2H_2O$, where m may be ≥ 1 and in a particular case $=1$. In the formula only H_2O+ is taken into account.

From a general viewpoint we have to emphasize that in all the kerolites studied the nature of the water bonds in the molecule is the same as in serpentines, since common (to serpentines and kerolites) endothermic (600—700°) and exothermic effects are regularly fixed on the thermograms. A roentgenometric study of these minerals shows that there are typical «serpentine lenses» on the powdergrams of both serpentines and kerolites. All this leaves no further doubt as to the common features in the chemical constitution of these minerals. An examination of the formulae of the members of the hydrophorsterite-kerolite series permits us to conclude that the chemical constitution of all these minerals could be interpreted by assuming that each member of this series (excluding the extremes) is a combination of two initial molecules: a molecule of the hydrophorsterite type and a molecule of the karachaite type. We should get simple stoichiometric ratios between the molecules mentioned in the minerals of the series. Thus, for instance, if we take the ratio $Mg_2SiO_4 : MgSiO_3$, we obtain the following values for these ratios respectively: № 1—1 : 0; № 2—3 : 1; № 3—2 : 1; № 4—1 : 1; № 5—1 : 2; № 6—1 : 2; № 7—1 : 3; № 8—1 : 0; № 9—1 : 0.

We consider the mineral species mentioned to be most stable and simple stoichiometric combinations of the above two initial molecules discovered so far. Doubtlessly intermediate members between them exist in nature. The existence of the orthosilicate part in the serpentine molecule ([3]) cannot be doubted, since it has been definitely proved through experiments, in which various serpentines were treated with NH_4Cl, and which were conducted in collaboration with engineer M. A. Oganesjan in the Rostov Institute of Applied Chemistry. On this treatment two thirds of the magnesium in the serpentine are extracted in the form of $MgCl_2$. Finally, it is of great interest to note that there exist in nature nickel, ferrous (ferri and ferro) and aluminium analogues for all of magnesial hydrosilicates examined by us. Thus, for instance, the existence of a new serpentine mineral kolskite, described in this article, seems to arise as a natural inference from the fact that there is a mineral, which exists in nature and has been described([11]) under the name of «ferrigalluasite-garnierite», which contains 29.4% NiO, 1.6% MgO and 14.42% $Al_2O_3+Fe_2O_3$, and which is an intermediate member between the pure nickel analogue of kolskite—$5NiO.4SiO_2.nH_2O$ and the corresponding alumo-ferri analogue—$5R_2O_3.12SiO_2.nH_2O$. It is thus a very simple matter to explain the chemical composition of the manifold sialites, containing Mg, and magnesial hydrosilicates, containing Al, if we assume that MgO is substituted with Al_2O_3. Thus, we are inclined to consider all these mineral formations—mineral «hybrids», members of definite series, whose initial molecules are magnesial hydrosilicates, on the one hand, and their aluminium analogues, on the other([10]). When substitution takes place according to the type $3MgO \rightleftarrows Al_2O_3$* various sialites of the baidellite, nontronite, bowlingite, saponite, volkonscoite, revdinite and other types are generally formed ([9]). In explaining the origin of the various minerals mentioned above through the combination of initial molecules of magnesial hydrosilicates and their aluminium analogues, we believe that we have here to do with a phenomenon which may be reduced to solid solutions. In this connexion we can now speak not only of analogy (as we understand it) in the chemical formulae of the initial molecules, but also of a certain geometric analogy in the lattices of the minerals in question, which vary widely in chemical composition.

All this goes to confirm the deep-lying relationship between magnesial silicates and their aluminium analogues, which had been pointed out by

* Or their partial isomorphous substitutes Fe″, Ni, Mn and F‴, Cr etc. respectively.

A. E. Fersman a long time ago, and which is rooted in the structure of the lattices of these minerals. It is probably by dint of this kinship that the initial molecules mentioned, substituting one another, give so readily a great variety of minerals of the intermediate type, which had been formerly taken for krypto-crystalline mixture.

The regularity of these phenomena, which is based on vast experimental material, is so wonderful that in our opinion it should draw intent attention, since profound work on this problem will doubtlessly be a turning point in our views on the nature of a very great number of mineral bodies heretofore considered complex and intricate.

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METEORITICS

FALL OF THE STONY METEORITE ZHOVTNEVY

By E. L. KRINOV

(Communicated by W. I. Vernadsky, Member of the Academy, 17. I. 1939)

The meteorite fell near the farm of Zhovtnevy, in the Stalin District of the Ukrainian Soviet Socialist Republic ($\varphi = 47°35'$ and $\lambda = 6°55'$ from Pulkovo) at 2 a. m. ± 10 minutes (time of 2nd zone), October 9, 1938*.

The fall of the meteorite was preceded by strong light and sound phenomena. At first the bolide had the appearance of a bright star (as bright as Venus), then it rapidly grew in brightness until it reached the visible dimensions of a full moon. The bolide made its path across the sky in about 5 seconds, moving approximately from N. W. to S. E., evidently at a relatively small angle with the surface of the earth. The bolide was ball-shaped and had a small tail. Shortly before its disappearance two pieces broke loose from it and flew into different directions, their radiance rapidly vanishing. While it was disappearing, the bolide shot sparks in all directions. Some of those who watched it observed sparks even while the bolide was in motion; Levkovsky, one of the spectators, heard some faint cracking. The locality was lit up so brightly that all objects in the houses could be seen quite clearly.

A few seconds after the disappearance of the bolide three or four blows of equal loudness were heard following each other without any interruption. The blows caused the window-panes to shake, some of the eye-witnesses even starting from their sleep after the first blow. The blows were followed by a great noise resembling that of a «gale», «a shower», «the flight of a flock of rooks», and so on. All this created panic not only among people, but also among domestic animals («the bulls bellowed»). Klipenko, Svetlichny, Temnokhud, eye-witnesses from the farmstead Zhovtnevy, have noted the direction of the noise and heard the thud of hard objects falling after the noise had ceased. The noise was also heard in the village of Prechistovka, 6 or 7 km N. W.**

After the fall of the meteorite four specimens of the latter (Fig. 2) were found in the neighbourhood of the Zhovtnevy farm (Fig. 1); probably many more of them fell down.

The first and evidently largest specimen which originally must have weighed about 32 kg was incidentally discovered by M. D. Bushna, a collective farm woman from Zhovtnevy, on October 11, last year, when she was in the field harvesting corn. The meteorite was lying in a hole, its

* The date October 10 which appeared in the press is erroneous.

** At present all the collected material is being worked up. Of interest is the coincidence of dates: the fall of the meteorite and the shooting star shower (on October 9, 1933), and also the close proximity of their radiants.

bottom part at a depth of 70 cm and its upper part at a depth of 20 cm. A few pieces were broken off by collective farmers immediately after the meteorite wos found. The following day teacher Borischenko extracted the meteorite from the hole and brought it to Zhovtnevy. When the meteorite was been extracted, it was discovered that it had been split in two unequal parts, one amounting to about one third, the other to about two thirds of the whole.

The hole, 70 cm deep, had a diameter of about 50 cm at the mouth. The bottom of the hole was of an elongated form corresponding to the shape of the meteorite. The diameter of the bottom of the hole was 33 cm from N. to S. and 11.5 cm from E. to W.

Both parts of the meteorite, particularly the smaller one, were broken into many pieces by the local people. The author succeeded in collecting 46 of these pieces, their total weight amounting to 6252.05 g (Fig. 2)*. Two of the largest pieces, forming the bulk of the specimen and weighing together 25 kg, were received by the Mariupol Museum at the beginning of November. Probably more than 10 pieces totalling a few kilograms in weight had been retained by local people and some others destroyed out of idle curiosity.

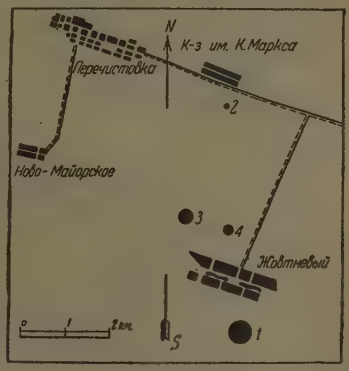

Fig. 1. Schematic plan of locality where the Zhovtnevy meteorite fell. The circles indicate the place where the particular specimens of the meteorite were found. The numbers of the specimens are marked by the side of the circles. The circles are of different size, indicating difference in the size of the meteorites.

Shortly after the discovery of the first meteorite the second was found by N. M. Friz, a shepherd of the Karl Marx Collective Farm. It was round-shaped and fractured on one side; originally it must have weighed about half a kilogram. This meteorite too was broken into many pieces. The author received two of them which together weighed 306.89 g (Fig. 2).

The third meteorite was found by P. K. Matveyenko, a pupil of the Zhovtnevy school. The meteorite was lying in a little saucer-shaped hole formed by it. Evidently when the meteorite struck against the earth about a fifth of it broke off. At the place where the meteorite fell more than 20 little pieces were found. Altogether the author obtained 22 pieces totalling 18 462.62 g in weight, two of the largest pieces weighing respectively 14 953 and 3174 g (Fig. 2).

The fourth meteorite was found by J. A. Shelest, a shepherd of the Zhovtnevy farm, about the same time when the first meteorite was found. The meteorite lay embedded in a field just dug up after the beet harvest, only one third of its bulk stricking out above ground. It was a piece, 2736 g in weight (Fig. 2), which evidently broke loose from the main mass in the air. The main mass probably struck earth somewhere near by, but has not been discovered yet.

Besides these pieces the Meteorite Committee of the Academy of Sciences received eight more by parcel post from D. F. Chechet. These pieces have probably been broken off from the first meteorite. Their weight totalled

* After this article has been sent to press, the Committee for Meteorites received some more pieces of the first meteorite specimen; from Falchenko—1 piece weighing 2646 g, and from Zayatz—3 pieces with total weight 251.55 g. Thus, the total weight of all the pieces of the first specimen at the disposal of the collection of the Academy of Sciences makes up 9149.60 g.

86.1 g. The total weight of all the pieces of the Zhovtnevy meteorite contained in the meteorite collection of the Academy of Sciences of the USSR thus amounts to 30 741.21 g.

The Zhovtnevy meteorite is the fourth of those found on the territory of the USSR in 1938 and the 93rd of all meteorites registered in the USSR. It is a typical grey chondrite containing a large number of chondrules, sometimes of a considerably large size (a few mm). Where the meteorite is fra-

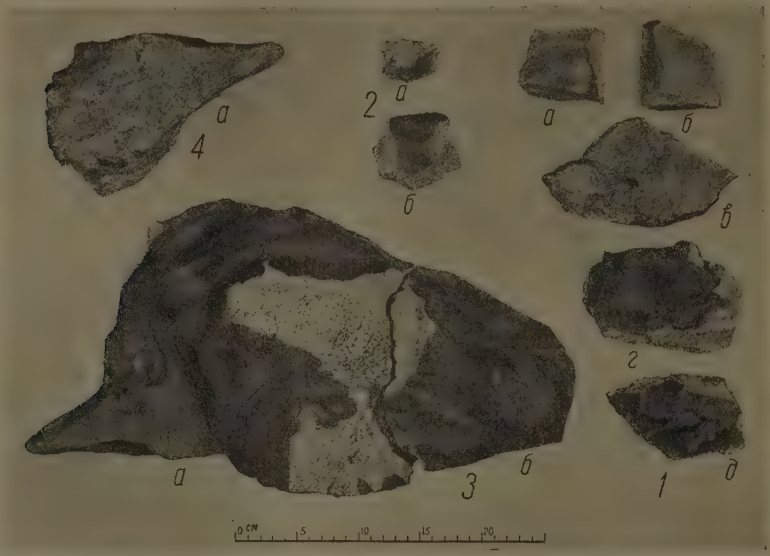

Fig. 2. The Zhovtnevy meteorite. The figures indicate the sequence in which the various specimens were found. The particular broken pieces are marked by letters.

ctured, numerous shiny metallic inclusions (nickeline iron, troilite and others) may be seen. In the fractured parts of the 2nd and 4th meteorites these inclusions in the form of thin laminae or knots attain a few centimeters in size. Very seldom small thin filaments and white inclusions resembling chalk are to be observed in the meteorite. The dark grey lusterless crust of the meteorite is thickly covered with piezoglypts. Two planes of the third meteorite show that the latter had evidently broken loose from the original mass not long before it touched the ground. These planes are covered with small rounded piezoglypts; on the other hand, the piezoglypts on the remaining surface are considerably larger, up to 5 cm in size; they are more elongated and deeper. The same may be observed on the fragment of the fourth meteorite. The meteorite is fairly brittle; this may be noticed directly and is moreover evidenced by the fact that all the four meteorites had split when falling, and some of the pieces show cracks.

The examination of the circumstances of the fall was made in November and December 1938 by the author of these notes who at the same time collected the pieces of the described meteorite.

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GENETICS

SOME STRUCTURAL DETAILS OF *CREPIS CAPILLARIS* CHROMOSOMES

By N. T. KAKHIDZE

(Communicated by N. I. Vavilov, Member of the Academy, 17. I. 1939)

In studying chromosome morphology there is frequently observed, in addition to the primary articulation which is always present, a further differentiation of the body of the chromosome. First, there are secondary articulations in the form of satellites or appendages on certain chromosomes; these are known to be of functional significance in the ontogeny of the nucleus, viz., in the process of formation of the nucleolus. Among secondary articulations are also included divisions of an arm of a chromosome into two or three segments, which are distinguished from appendages by the fact that in most cases they are united by a common peripheral layer of chromatin («sheath»), which makes it a more difficult problem of technique to bring them out distinctly (1). As a means of bringing them out more clearly there has recently come into use a method of fixation with solutions containing no or almost no acetic acid, sometimes preceded by cooling of the root-tips.

In the present investigation such a procedure, combined with the method of germination in sawdust used for onions (to bring out the chromomere structure of the chromosomes), was utilized for *Crepis capillaris*, a plant widely used in cytological studies.

The precise structure of the chromosomes of *Crepis capillaris*, as given by a count of the number of chromomeres in chromosomes *A* and *C*, has been described by Geitler (2), who used a modification of Flemming's solution to fix his material. The writer also succeeded to some extent in differentiating the chromosomes into chromomeres, but their detailed description will be deferred until the chromomeres have been more completely individualized.

The results given below were obtained with the following fixatives:

1) Modification of Flemming's solution (after Geitler).

2) Chromic formalin: 1% chromic acid—5 cm^3+40%formalin—5 cm^3.

3) Modification of Champy's solution: 1% chromic acid—$3^1/_2$ cm^3; 3% bichromate of potassium—$3^1/_2$ cm^3; 2% osmic acid—$1^1/_2$ cm^3; glacial acetic acid—2 drops.

In order to bring out the secondary articulations, a combination of all three conditions (sawdust, pre-cooling of root-tips, special fixative) is apparently necessary, since otherwise they are not visible in *Crepis*.

In the first vertical row of the accompanying figure are shown the three chromosomes of the set when fixed with chromic formalin without pre-

cooling of the root-tips (A_1, C_1, D_1). By using the special method described above the following additional articulations were revealed: in chromosome A the short arm is differentiated into two unequal segments (A_2); besides this, there is often seen quite a marked constriction in the long arm (A_3). In the long arm of chromosome C there is a secondary articulation, setting off approximately one-third of this arm (C_2). Chromosome D also has an articulation in the long arm (D_2).

As regards all these articulations, it should be noted that they are far from being observed in all cells. This is a common phenomenon, observed also in other plants, and is presumably connected with the condition of the cell and its position in the root-tip. But in those nuclei, where these articulations are observed, they are located at quite definite loci. When the root-tips are pre-cooled and the chromomeres well brought out, the secondary articulations are still visible and stand out as distinctly as ever (A_4). If the root-tips are not pre-cooled, the body of the chromosome is rather uniformly divided into chromomeres (A_5). This may be seen also from the figures and photograph in Geitler's paper. The body of a chromosome, although consisting of a more or less uniform chain of chromomeres, seems to contract unevenly when subjected to cooling. It is not difficult to find similar data in literature. One of the best examples is that of *Secale cereale*, in which a considerable number of secondary articulations were revealed, when the method of pre-cooling the root-tips was used ([3]). In investigations without pre-cooling of the root-tips, despite exceptional care in analysis, such articulations were not revealed ([1]).

Notwithstanding their dependence upon the method employed, data on articulations, inasmuch as they are specific, constitute a morphological characteristic of chromosomes and show the specifity of their structure. Hence, they may play the practical rôle of specific markers in comparative analyses of idiograms of different species and in the identification of separate elements in structural chromosome changes. It may prove possible to find a connexion between the revealed differentiation in the bodies of *Crepis capillaris* chromosomes and the points of rupture in the chromosomes of this plant when subjected to X-radiation. It would be of particular interest to do this for chromosome D, where investigators have most frequently found inversions and translocations, in which the proximal end of the chromosomes is detached at a point very close to the secondary constriction ([4]). It is possible that the loci of such secondary constrictions are the most favourable (though, of course, not the only) points for a break in the continuous chain of chromomeres.

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GENETICS

ACTIVE CONCENTRATIONS OF ACENAPHTHENE INDUCING ALTERATIONS IN THE PROCESSES OF CELL DIVISION IN PLANTS

By A. SHMUCK and A. GUSSEVA

(Communicated by N. I. Vavilov, Member of the Academy, 20. I. 1939)

The characteristic action of acenaphthene upon the processes of cell division in plants leading to various grades of polyploids was established by one of us [1] and confirmed by a series of works, both with regard to cytological changes taking place in cells under the influence of acenaphthene [2,3,4] and to production of fertile tetraploid plants[5].

Further on it was established that the peculiar action of acenaphthene is produced by sublimated particles [6] and that comparatively very small quantities of the latter may effect very significant disturbances of the processes of cell division in plants.

Until recently it remained unclear what quantities of acenaphthene are necessary for the production of a corresponding biological effect; in the present work we have introduced precision in this question, the solution of which was necessary also for comparison of the strength of action of individual substances applied for the production of polyploid plants.

The investigation has been conducted on summer wheat; yet, the standards worked out by us are applicable to many other cultivated plants of the grass family.

Owing to the fact that it was established for a number of carcinogenous substances that the effect observed may be produced not by the substance itself but by admixtures contained in it [7], it was necessary to be sure in the case of acenaphthene that the characteristic action of acenaphthene upon plants actually depends upon this very substance and not on some admixtures contained in it.

Therefore, we have subjected acenaphthene to purification by means of repeated recrystallization from ethyl alcohol; the mother-liquor was collected and impure acenaphthene contained in it was separated by means of evaporation of the alcohol.

Further purification consisted in the following. From repeatedly recrystallized acenaphthene was prepared a compound with picric acid. Picrate of acenaphthene was obtained from a reaction between acenaphthene and picric acid in alcoholic solutions in the form of a crystalline precipitate of a reddish-brown colour which had a very definite melting point at 161°.

The picrate after purification by two crystallizations from alcohol was broken up by aqueous alkali, acenaphthene was precipitated, collected by filtration and two more times recrystallized from alcohol.

Pure acenaphthene was prepared in this way and comparative experiments were arranged on its action upon germinating seeds of wheat.

For sake of comparison were tried common commercial acenaphthene, acenaphthene purified by six crystallizations from alcohol, impure acenaphthene obtained from the first mother-liquid and pure acenaphthene obtained from the picrate.

For the trials ether solutions were prepared; the acenaphthene content of these solutions was exactly determined and definite quantities of them were put into Petri dishes (of 10 cm diameter) directly on filtrating paper; after evaporation of the ether, the paper was moistened with water, the amount of which was equal in all the dishes, and 10 kernels of wheat were placed into each dish.

It was revealed by this experiment that the biological effect of acenaphthene, as manifested by tumor-like thickening of the sprouts, shortening of the roots and lanceolate swelling of the root-tips, is expressed even in the case of the smallest dosages of acenaphthene. Within the limits from 0.1 to 0.001 g of acenaphthene its action in all the cases appeared to be absolutely identical in its strength.

When smaller quantities of acenaphthene (parts of a milligram) were introduced, it proved to be possible to establish with sufficient exactitude the limit of concentration when the active influence of acenaphthene ceases altogether.

For wheat such a limit of action of acenaphthene proved to be 0.2 mg per 10 kernels in a Petri dish. From this dosage up we always obtained similar effect as shown by shortening of sprouts and roots, by significant tumour-like thickening of the sprouts which is characteristic of the action of acenaphthene; the dosage below 0.2 mg per Petri dish produced no influence upon the sprouts of wheat.

The dosages of acenaphthene which were up to the present day applied for the production of polyploid plants are far too high and may be replaced by much lower ones (not more than 1.0—1.5 mg per Petri dish) which not only leads to saving of acenaphthene, but also to production of healthier plants, since the roots of the seedlings always suffer from the excess of acenaphthene. The application of acenaphthene in the form of ether solution is very convenient.

After having established in the above way the lower limit of the concentration of acenaphthene when its biological effect is clearly expressed, we could compare with each other the strength of action of the acenaphthene preparations obtained by us.

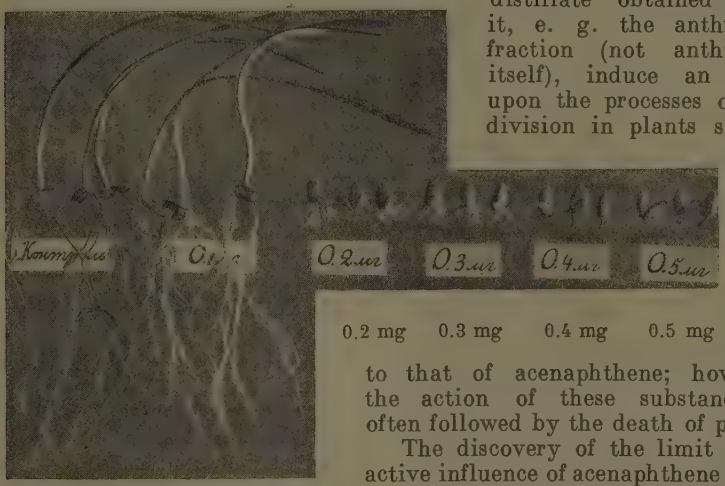
	Quantities of acenaphthene per Petri dish				
	0.5 mg	0.4 mg	0.3 mg	0.2 mg	0.1 mg
Commercial acenaphthene . .	Strong action	Strong action	Strong action	Limit of action	No action
Acenaphthene crystallized six times	Do	Do	Do	Do	Do
Acenaphthene from the 1st mother-liquor	»	»	»	»	»
Acenaphthene prepared from the picrate	»	»	»	»	»

These experiments convinced us that all the preparations of acenaphthene possess absolutely equal biological activity; the characteristic action of

acenaphthene upon plants leading to polyploids depends, therefore, on this very substance and on no eventual admixtures.

It is of interest to note that the picrate of acenaphthene produces an effect which closely resembles the characteristic action of acenaphthene, while picric acid either causes death of the plant or, in weak concentration, has no effect whatever.

Owing to the fact that acenaphthene is contained in the coal-tar (from which it is prepared), the coal-tar itself as well as individual fractions of the distillate obtained from it, e. g. the anthracene fraction (not anthracene itself), induce an effect upon the processes of cell division in plants similar



Control 0.1 mg 0.2 mg 0.3 mg 0.4 mg 0.5 mg

to that of acenaphthene; however, the action of these substances is often followed by the death of plants.

The discovery of the limit of the active influence of acenaphthene which is at 0.2 mg per 10 germinating seeds of wheat enabled us to arrange further tests in order to determine the amount of acenaphthene sufficient to produce effect on 1 germinating seed of wheat. To do this we increased the number of kernels per Petri dish, while the total dosage remained the same (0.3 mg).

Number of kernels per Petri dish	10	20	30	40	50	80	100
Effect	Strong	Strong	Strong	Strong	Weak	Nil	Nil

Two tenth of a mg of acenaphthene is thus sufficient to cause a striking effect which is manifested in disturbances in cell division in 40 kernels of wheat. Under these conditions each kernel cannot absorb more than 0.007 mg of acenaphthene.

An extremely low quantity, less than 7 micrograms, of acenaphthene represents a biologically active dosage of this substance which disturbs cell divisions in 1 kernel of wheat and may induce formation of polyploid chromosome sets.

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GENETICS

OBSERVATIONS IN VIVO ON THE FORMATION OF THE MALE SEX CELLS IN *LILIUM MARTAGON* L.

By X. J. KOSTRIUKOVA

(Communicated by N. I. Vavilov, Member of the Academy, 24. I. 1939)

Lilium martagon had been many times studied by the most outstanding cyto-embryologists of plants ([1], [2], [3], [4], [5], [6], [7]). With the appearance of S. G. Navashin's work, regarded as classical in the world literature, there was held established that the division of the generative cell results in the formation of male gametes which are naked nuclei. In this work Navashin left unsolved the question as to the small bodies observed in the cytoplasm of the generative cells; he suggested the idea that they were some remainder of the blepharoplast.

In summer 1937, as soon as it was made clear, owing to the author's preceding work, that the living, moving and dividing generative cells are directly accessible to observation under high power, at least in some *Amaryllidaceae*, there was undertaken, according to W. W. Finn's advice, an investigation in vivo of the well-known object of cytological observations—*Lilium martagon*. With the beginning of the blooming of *Lilium martagon* the pollen of this plant was sown on various artificial media which are known from literary data and from the author's personal observations. However, though the pollen of the lily was germinating on various media, the pollen tubes died earlier or later, before the division of the generative cell was accomplished.

Not before 1938, on germinating pollen grains in Boetcher's chamber in a fluid secreted by the stigma of *Lilium longiflorum*, *L. candidum* and *L. regale* in 10% sugar solution, I have succeeded to realize favourable conditions for the development of the male gametophyte of *L. martagon* and to trace this development directly in vivo.

In living unstained pollen grains of *L. martagon* the generative cell is not seen. The pollen tubes begin to grow approximately an hour after sowing. The content of the tubes has the appearance of an emulsion consisting of droplets of various size rapidly removed by the flow of the plasm.

In spite of the poor transparency of the plasm of the pollen tube, the generative cell in it can be rather often well seen. If it is directly underlying the wall of the pollen tube directed toward the objective of the microscope, then odservation is not prevented by the flows of plasm the tube is washed with. Being large and occupying almost the entire lumen of the pollen tube, the generative cell may rather often be seen in such a position. Amidst the plasm it is surrounded by it has the appearance of a light elongated body.

Sometimes one succeeds in discerning its posterior rounded and anterior stretched ends. At the poles of the generative cell bodies are seen of pale greenish colour. At this stage of development the body in the posterior end is rounded, while that in the anterior end is somewhat elongated, according to the stretched form of the anterior end.

Somewhat later there appear in the nucleus structures of another light-refractivity, the chromatin threads. At the beginning the chromatin threads form a complex interlacement and appear in one optical plane as bent short segments moving into various directions. Fig. 1 represents the described stage—the prophase of the division of the generative cell.

Further the generative cell stretches, its nucleus grows, becoming at the same time lengthened. The chromatin threads, now chromosomes, are arranged chiefly along the long axis of the stretched nucleus. Sometimes one succeeds in discerning that the transparent threads, chromosomes, have a symmetrical arrangement in the nucleus. The coloured bodies lie also symmetrically on both poles and the very shape of the cell becomes symmetrical. Observations in vivo previously effected on processes of division of the generative cell in *Amaryllidaceae* permit of defining the stage as that of metaphase (Fig. 2). It is to be noted that among all the plants studied by me in vivo (13 species of *Liliaceae* and 8 species of *Amaryllidaceae*) *L. martagon* is the one in which these structures are most distinct. However, owing to the movement of the generative cell and of the pollen tube plasm, as well as to the change of these structures themselves and their transparency, one fails to view them in detail, just as one fails to trace the bendings of one and the same chromosome, as we use to do on fixed and stained material.

In the next stage of development the generative cell is still elongating. The transparent threads, chromosomes, are arranged along the long axis of the strongly stretched cell in parallel rows. At this stage no transverse threads or bendings are seen, which are observed in the middle part of the cell at the preceding stage. The coloured bodies are divided into regions and are seen not only at the poles, but also in the middle part of the cell. Fig. 3 represents this stage—the anaphase of the division of the generative cell.

Fig. 4 represents the generative cell of lily at the stage of telophase. The chromatin threads form at this stage filial spirems. The coloured bodies are divided into regions lying also in the transparent middle part of the cell.

The division of the generative cell ends in the formation of well formed male sex cells rich in plasm (Fig. 5 and 6). At early stages of formation of sperms the nuclei of the cells are rounded and chromatin structures are seen in them. Further on the nuclei lengthen considerably, a cell wall-like formation is laid between the sperms, which is seen as a well distinct light stripe of a light-refractivity different from that of the cytoplasm of the cell. The coloured bodies are arranged at both poles of the filial nuclei. When the outlines of the nuclei lack distinctness, one can determine the form of the nuclei by the form of coloured bodies the nuclei seems to be stream-lined with (Fig. 5). Even at this stage of development of the male gametophyte the coloured bodies are completely distinct, but their coloration is often paler than at the stage of the generative cell.

When fixing pollen tubes of *L. martagon* with the chondriosome fixing fluid with osmic acid after G. A. Levitsky, I have succeeded in determining also on fixed material the presence of sperm-cells rich in plasm. As to the cell wall-like formations between the sperms, on the one hand, and the coloured bodies, on the other hand, all of them are poorly preserved in fixed material, and seem to be dissolved even by such a delicate fixing fluid which permits to preserve the finest plasmatic structures. Sometimes at poles of the cells, in the places of disposition of the coloured bodies, small regions are

seen which become black under the action of osmic acid. It should be supposed that these are the remainders of the coloured bodies destroyed by the fixing fluid, the more that the blackened areas sometimes keep the general shape of these bodies (distended towards the nucleus). Navashin probably described these very bodies in the plasm of the generative cell of the lily as «some remainder of the blepharoplast».

Thus, only the investigation in vivo permits to elucidate the structure of the coloured bodies. As early as at the beginning of the investigation I succeeded in establishing that the bodies are no compact formations. I believed them to consist of grains or droplets close-packed on a small area at the

poles of the nuclei. Such a structure of these bodies seemed the more probable that they were stained during their life-time by Neutralred, as granules in the cytoplasm of the generative cell of *Amaryllidaceae*. However, when observing these bodies in vivo I was not sure that such an interpretation of their structure was true. Only in summer 1938 I have succeeded to observe in especially successful and extraordinarily transparent cultures, with large magnifications of the microscope (×1500), winding interlaced canaliculi, nuddled together in a clot which I gave the name of «a coloured body». Fig. 7 represents the end of the generative cell with a coloured body of a round shape, as it is often to be observed soon after the passage of the generative cell into the pollen tube. The winding canaliculi are shown as they are seen in one optical plane. On rotating the micrometer screw one can become persuaded that while bending they rise or sink forming a complex compact interlacement.

The delicate structure of the coloured bodies, their reaction to osmic acid, their behaviour during cell division strikingly remind of the formation discovered by Golgi in 1898 in the nerve-cells and named by him «apparato reticulare interno», and recently interpreted by Guillermond (8), at least for plants, as vacuole at various stages of development of the cell.

It seems to me most likely that these extremely delicate structures described by me, which can be discerned in vivo only by an expert eye owing to their natural colour, correspond to the structures described by many authors in animal and plant cells under the name of Golgi's reticular apparatus. This is also corroborated by the relation of these structures to the staining by Neutralred in vivo.

The classical work of S. G. Navashin, describing the sperms in *L. martagon* as naked nuclei, is up to now made use of as a serious argument in support of the theory of the monopoly of the nucleus in heredity-phenomena (Course of Genetics by Grischko and Delaunay, chapter 1, 1938).

All the figures were made with Abbe's drawing-apparatus with the magnification: ob. 100 Zeiss × oc. 5, except Fig. 7—ob. 100 Zeiss × oc. 10. Explanations in text.

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HYDROBIOLOGY

ORGANIC MATTER IN THE WATERS OF THE BARENTZ, POLAR AND KARA SEAS

By B. A. SKOPINZEV

(Communicated by S. A. Zernov, Member of the Academy, 15. I. 1939)

The lack of data on the content of organic matter in the waters of our northern seas induced us to undertake appropriate observations during the 40th cruise of the research ship «Perseus» (August-September 1932).

The observations were carried out in the Barentz, Polar and Kara seas. All the determinations were made by the method of «oxidability» (in a neutral medium)*. This method is as far the only one applicable under expeditionary conditions.

The mean results of two parallel determinations in mg O_2 per litre are given in Table 1. In the figure are plotted the stations where the determinations of oxidability were made.

These data show that the oxidability value is subject to great variations, with a minimum in the northern part of the Barentz sea and a maximum in the Kara sea. This conclusion is based on the comparison of the oxidability values of water samples, taken at levels of 25, 50 and 100 m. As to the oxidability value of the surface layer (0 m), we are not certain that it is characteristic of the organic matter of the sea water alone, since contamination from the boardside is always possible in this case.

A more detailed analysis of the material obtained permits of several generalizations. To that effect the mean values for each level of all the three seas are given in Table 2.

The calculated mean values of this table point out that:

1. The oxidability decreases with the depth, the bottom watter being characterized in some cases by a somewhat higher oxidability.

2. The Kara sea has the highest oxidability value of the surface layer; this agrees with the presence of a powerful continental run-off in the sea.

3. Among all the stations taken in the Barentz sea we can clearly distinguish those with the lowest oxidability values ($\leqslant$0.30 mg O_2 per litre) on

* The details of the method are as follows: 10 cm³ of sea water are sampled in tubes with glass stoppers; then 0.3 cm³ of 0.01 standard solution of $KMnO_4$ are added and the tubes immersed in a boiling bath for 20 min. Upon cooling KI is introduced, H_2SO_4 added and the iodine produced is titrated with a 0.001 standard solution of hyposulphite. All the determinations were duplicated, the maximum deviations from the mean being 5%. The comparison of this method and the usual method ([1],[2])—100 cm³ of water+10 cm³ of 0.01 standard solution of $KMnO_4$ with subsequent boiling for 10 min.—confirmed their practical agreement. We were obliged to introduce such a modification, the usual method not being for use on board ship.

Table 1

Results of the Determinations of Oxidability of Sea Water (neutral medium) in mg O_2 per litre

No of stations	0 m		25 m		50 m		100 m		Bottom		
	pH	Oxidability	pH	Oxidability	pH	Oxidability	pH	Oxidability	Depth	pH	Oxidability
					I. Barentz sea						
2152	8.21	0.42	8.19	0.42	8.15	0.34	8.06	0.21	250	8.06	0.25
2157	8.17	0.58	8.17	0.46	8.13	0.34	—	—	102	8.10	0.36
2160	8.19	0.60	8.19	0.27	8.15	0.32	—	0.33	166	8.00	0.26
2174	8.18	0.58	8.17	0.30	8.16	0.30	8.14	0.19	223	8.11	0.25
2175	8.17	0.44	8.18	0.26	8.17	0.24	—	—	111	8.14	0.22
2182	8.16	0.63	8.16	0.38	8.16	0.33	8.13	0.38	335	8.10	0.30
2184	8.18	0.52	8.19	0.44	—	—	8.11	0.30	248	8.09	0.36
2186	8.23	0.51	8.21	0.42	8.19	0.27	8.15	0.25	272	8.11	0.30
2187	8.22	0.59	8.23	0.32	8.20	0.29	—	—	128	8.18	0.25
2189	8.18	0.34	8.18	0.23	8.17	0.23	8.22	0.20	308	8.06	0.21
2190	8.18	0.34	8.18	0.23	8.17	0.20	8.13	0.29	322	8.08	0.34
2191	8.18	0.54	8.18	0.28	8.16	0.27	8.11	0.20	309	8.06	0.38
2192	8.17	0.43	8.18	0.30	8.15	0.27	8.12	0.27	339	8.08	0.35
					II. Polar sea						
2162	8.21	0.53	8.21	0.34	—	—	—	—	53	8.11	0.30
2163	8.18	0.58	8.20	0.42	8.17	0.30	—	—	102	8.11	0.35
2168	8.11	0.47	8.13	0.31	8.18	0.34	—	—	135	8.14	0.26
2169	8.17	0.50	8.16	0.51	8.15	0.46	8.15	0.34	208	8.07	0.34
					III. Kara sea						
2196	8.19	0.82	8.19	0.46	8.17	0.30	8.18	0.31	290	—	—
2198	8.12	0.75	8.10	0.37	8.09	0.33	8.08	0.46	184	—	—
2200	8.15	0.78	—	—	—	—	—	—	30	8.13	0.42
2202	8.12	0.60	—	—	—	—	—	—	18	8.11	0.50
2203	8.11	0.69	—	—	—	—	—	—	29	8.10	0.46
2204	8.12	0.76	8.14	0.46	8.13	0.34	—	—	127	8.11	0.26
2210	8.15	0.60	8.15	0.42	8.19	0.50	—	—	101	8.10	0.42
2211	8.16	0.54	8.15	0.45	8.11	0.45	8.09	0.37	180	8.09	0.42
2214	8.16	0.42	8.14	0.34	8.10	0.31	—	—	140	8.08	0.25
2216	8.14	0.54	8.14	0.38	8.13	0.34	—	—	103	8.10	0.30
2218	8.14	0.47	8.16	0.38	8.14	0.38	—	—	206	8.10	0.30
2220	8.12	1.10	—	—	—	—	—	—	17	8.12	0.61

depths of 25 and 50 m. Other stations have values higher than 0.30 mg O_2 per litre on the same levels.

Comparing oxidability values and water temperatures, we see (with a few exceptions*) that a low oxidability is characteristic of waters with low temperatures (from + 1.0 to—1.6°), while waters with high temperatures (from + 4.5 to + 1.0°) are characterized by high oxidability values. It should be mentioned in this connexion that high temperature waters of the west and north-west parts of the Barentz sea are «fresh» Atlantic waters, whereas the low temperatures of the northern and north-eastern waters indicate their genetic relation to different areas of the Polar sea.

Thus, our data show that the waters of the west and north-western part of the Barentz sea contain more organic matter, than those of the northern

* The exceptions refer to stations situated south and east from Spitzbergen, where a specific influence of the continent may be assumed,

Table 2

Mean Values of Oxidability of Sea Water at Different Levels in mg O_2 per litre

Depth in m	Barentz sea				Polar sea		Kara sea	
	Stations No * 2152, 2157, 2182, 2184, 2186		Stations No** 2160, 2174, 2175, 2187, 2189, 2190, 2191, 2192					
	Extreme values	Mean values	Extreme values	Mean values	Extreme values	Mean values	Extreme values	Mean values
0	0.42—0.63	0.54	0.34—0.60	0.46	0.47—0.58	0.52	0.42—1.10	0.67
25	0.38—0·46	0.42	0.23—0·32	0.27	0.31—0.51	0.40	0.34—0.46	0.41
50	0·27—0.34	0.32	0.20—0.32	0.26	0.30—0.46	0.37	0.30—0.50	0.37
100	0.21—0.38	0.28	0·19—0.33	0.25	0.34	0.34	0.31—0.46	0.38
Bottom < 100	—	—	—	—	0.30	0.30	0.42—0.61	0·50
Bottom 100—200	0.36	0.36	0.22—0.26	0.24	0.26—0.35	0.30	0.25—0.42	0.33
Bottom > 200	0.25—0.36	0.30	0.21—0.38	0.30	0.34	0.34	0.30	0.30

* Oxidability on levels of 25 and 50 m $>$ 0.30 mg O_2 per litre.
** Oxidability on levels of 25 and 50 m $<$ 0.30 mg O_2 per litre.

and north-eastern part*. The fact that this is not due to a greater development of phytoplankton in the «fresh» Atlantic waters during the investigation is testified by the practically identical content of dissolved oxygen and pH (Table 1) in all the stations, made in the Barentz sea (at the same levels).

What is then the explanation of the distribution of oxidability observed?

One of the outstanding features of the northern and north-eastern part of the Barentz sea is the floe-ice, covering the sea for the greatest part of the year. Under such conditions phytoplankton does not develop. In the western and north-western parts of the Barentz sea conditions are different, so that phytoplankton is developing almost all year round, except for the dark period. This fact is probably responsible for the greater accumulation of organic matter in the waters of that part of the Barentz sea (so much the more, as its consumption at the expense of oxidation processes proceeds practically at the same rate in all the areas mentioned).

Sure enough, to clear up the problem, further more extensive observations are needed, taking into account such factors as seasons, development of phytoplankton, distribution of floe-ice, etc.

We would note, moreover, that differences in oxidability values on levels of 25, 50 and 100 m are greater in waters with high oxidability, particularly in layers from 25 to 50 m.

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* Earlier investigations showed that the Barentz sea waters with low temperatures and reduced salinity are also characterized in a notable part by lower values of pH and dissolved oxygen(4).

In connexion with the above stated, of great interest is the particular character of the physico-chemical properties of the soil of the northern part of the Barentz sea (compared with the western part): those soils, brownish in colour and rich in MnO_2, belong to the group of oxidized soils (M. V. Klenova); their redox potential is relatively high—E_h=about 265 mV (S. V. Brujewicz), while they are characterized practically by total lack of chlorophyll (L. A. Yastrebova).

4. In the Polar sea, the area, where determinations of oxidability were made, is also traversed by warm currents (from +1.9 to +3.4°) (West Spitzbergen current); here the oxidability at depths of 25 and 50 m exceeds 0.30 mg per litre. The waters of this area are also characterized by a relatively low transparency (6.5—8.5 m) and high colour-numbers (Forel-scale) which, of course, influence to some extent the oxidability values (towards their increase).

⊙ Stations with oxidability ≤ 0.30 mg O_2 per litre at 25 and 50 m

● » » » > 0.30 » » » » » » » » »

5. A comparison of the values of Table 2 and those of the oxidability of the south-eastern part of the Kaspian sea(6) (determined by the same method in a neutral medium) shows that the latter is on the average 3—5 times higher, than the oxidability of the Barentz, Kara and Polar seas.

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PLANT PHYSIOLOGY

INFLUENCE OF β-INDOLYL-ACETIC ACID UPON PHOTOSYNTHESIS

By N. G. CHOLODNYJ, Member of the Academy of the Ukrainian SSR, and
A. G. GORBOVSKY

Heteroauxin of β-indolyl-acetic acid acquired wide reputation during the last years owing to its remarkable physiological activity. Its action upon the growth of cells of higher plants and upon various phenomena of morphogenesis is generally known. There are indications (rather inconsistent, it is true) that respiration, movement of the protoplasma and some other functions of the vegetable cell experience noticeable alterations under the influence of very weak solutions of heteroauxin. It was established by the recent studies by Mitchell and Hamner ([1]) that β-indolyl-acetic acid may under certain conditions increase accumulation of dry substance in leaves and in other organs of green plants. The conclusion is based by the authors upon experiments with young decapitated bean plants (*Phaseolus vulgaris*), in which lanoline paste containing heteroauxin was applied to the surface of the cut made through the upper part of the stem. Such plants after a few days were compared with the control plants, which were similarly decapitated but treated with lanoline paste containing no heteroauxin. It was found that experimental plants (treated with heteroauxin) within three days accumulated nearly twice as much dry material as the controls, in spite of the fact that both were placed under exactly identical conditions and possessed identical leaf surface.

Mitchell and Hamner do not dwell upon the question on the nearest causes of the interesting phenomenon established by them. The most probable explanation would be that heteroauxin contained in the paste, applied in their experiments to the surface of the cut, penetrated into the green leaves of the experimental plants and in some way stimulated their assimilatory activities.

In favour of this supposition speak experiments arranged by us in the summer and autumn of 1938 with the aim to find out what is the influence upon photosynthesis of weak solutions of β-indolyl acetic acid introduced into the laminae of isolated leaves. However, these experiments were designed and conducted without any connexion with the above-mentioned investigation by Mitchell and Hamner which came to our knowledge much later. The starting point of our experiments were the old data by V. N. Lubimenko and O. A. Stscheglova([2]) who studied the influence of wounding of lamella upon the intensity of photosynthesis. These authors notably arrived at a conclusion that mechanical injury of the living tissues of the leaf by its perforation causes augmentation of the energy of photosynthesis. This effect becomes apparent not at once; at the beginning a decrease in assimi-

lation is observed, and only after a certain time a rise of the photosynthetic curve begins, the maximum value being attained from 3, 5 to 6 days following the wounding of the leaf.

Lubimenko and Stscheglova gave no profound physiological analysis of the phenomenon described by them. They contented themselves by concluding that photosynthesis is a «process of irritability» and depends upon the physiological condition of the protoplasm, which is changed under the influence of most diverse external factors. It is natural, however, to raise the question as to what are the changes, which occur in the protoplasma of the cells of the leaf under the influence of mechanical injury, and how do these changes influence the process of photosynthesis.

We know since the time of the classical investigations of Haberlandt (1913—1923) that healing of wounds in higher plants is accompanied by accumulation in adjacent tissues of specific substances which possess properties of hormones.

These substances, whose chemical nature has not been cleared up as yet, in their physiological action upon the vegetable cell have much in common with the so-called growth hormones, the auxin and the heteroauxin.

If one takes into consideration that maximum rise of photosynthesis in injured leaves, according to Lubimenko and Stscheglova, coincides with the period of healing of the wounds and sometimes even follows it, the idea suggests itself that one and the same cause underlies these phenomena which at the first sight are so different from each other; it should be sought in the action upon the protoplasma of physiologically active substances which arise or are accumulated near the injured regions and then penetrate into the neighbouring not injured tissues of the leaf. These considerations have led us to the attempt to investigate the action of various phytohormones upon photosynthesis.

Leaves of lilac, poplar, jasmin, hemp, hydrangea and of other plants were taken for the experiments. Two leaves were selected on one and the same individual which were as far as possible identical with respect to age, size, position on the stem and conditions of illumination. These leaves were cut off (under water) and placed in two glass chambers of the apparatus for determination of intensity of assimilation as designed by E. F. Votchal ([3]). The ends of the petioles were immersed into Knop's solution. Ordinary air, not enriched with carbon dioxide, was passed through them with a speed of 500 cm^3 per minute. The upper sides of the leaves were illuminated with two identical lamps of 250 or 500 W each; each lamp was installed above one chamber at a distance of 45 cm. In order to prevent heating of the chambers, above them was installed a screen made of cardboard or asbestos and supplied with openings corresponding to the leaves; besides, the lamps were immersed during the whole experiment into half-circular baths with running water. In some of the experiments the upper surface of the chambers was also cooled by water.

For absorption of carbon dioxide absorbers of the system of A. A. Richter were used. As their absorbing capacity depends on the quality of the filters welded into the glass tubes, a set of absorbers was selected by means of preliminary experiments which gave identical values for the amount of carbon dioxide absorbed, when definite volumes of air were passed through them with equal speed. In these experiments the differences observed amounted only to tenths of a mg.

In each experiment with leaves three absorbers were used: two of them served to determine the amount of carbon dioxide in the air which passed through the chambers containing leaves, and the third to check the CO_2 content in the outer air.

The intensity of assimilation was expressed in milligrams of carbon dioxide decomposed during one hour by one square decimeter of the surface of the leaves.

The scheme of the majority of our experiments was the following. Two freshly cut leaves were put into chambers, and the amount of carbon dioxide decomposed by them during a 100—110 minute exposition was determined. Then the leaves were taken out, immersed with their petioles into Knop's solution and left in scattered light or in darkness for a few hours. A second determination of the amount of carbon dioxide decomposed by them within 100—110 min. was made under the same conditions. These two determinations made it possible to judge of the energy of photosynthesis of the leaves taken for the experiment.

After the second determination one of the leaves (the control one) was again immersed with its petiole into Knop's solution, and the other (the experimental one) in the same solution to which, however, a certain amount of β-indolyl-acetic acid was added. In the experiments of 1938 we have applied only two concentrations of this substance: 1 mg and 10 mg per one liter of Knop's solution. The leaves remained in the solutions for several hours, then they were replaced into the chambers for determination of the intensity of assimilation. After this third determination the leaves usually remained again for a more or less prolonged period of time out of the chambers and immersed into the same solutions and then were once more exposed in the chambers. Sometimes several such determinations were made in order to compare the photosynthetic energy of the experimental and the control leaves.

We shall descride in short two typical experiments conducted according to this scheme with weak solutions of heteroauxin (1:1000).

Experiment 1. Hemp. The foliate stem was cut off and put into pipe-water. After 7 hours two leaves were taken from it and immersed with their petioles into Knop's solution for 2.5 hours. Then the two leaves were placed into the chambers of the apparatus which was located far from the window (facing north-west) and illuminated from above by two lamps of 250 W each. The upper surface of the chambers was cooled with running water; the temperature of the air within them during all the determinations described in this experiment fluctuated from 23.6° to 24.6° C. Exposition was equal to 110 minutes.

The first determination showed that one of the leaves (No 1) had intensity of assimilation 10.00, the other (No 2)—12.10 (for the sake of brevity we shall further designate the corresponding values as i_1 and i_2). Then the two leaves were left for 2 hours in Knop's solution. The second determination gave the following values: i_1=9.6, i_2=12.9.

After this the 1st leaf was placed into Knop's solution with heteroauxin (1 : 1000), the 2nd—simply into Knop's solution. In these solutions the leaves remained throughout the night (9 hours) and then were placed into the chambers for 110 min. under the same illumination as before. This time i_1 proved to be equal to 19.24, i_2=14.73.

During the following four hours the leaves remained immersed into the same solutions. Then the intensity of assimilation was determined under the same conditions. The 1st leaf gave 18.0, the 2nd—13.8.

Thirty hours later, during which the leaves were left in the same solutions, the last determination was made: i_1=2.1, i_2=14.83.

Experiment 2. Hydrangea. Cut leaves were immersed with their petioles for 2.5 hours into Knop's solution, and then transferred to the chambers. For illumination were used lamps of 500 W. Exposition was equal to 107 min.; i_1=4.6, i_2=3.0.

The second determination was made 10 hours later, during which time the two leaves were in Knop's solution; 8 hours out of this time they were in darkness. Exposition 107 min. under the same condition of illumination and temperature; $i_1=5.1$, $i_2=3.34$.

After this the 1st leaf was immersed for 15.5 hours into Knop's solution with heteroauxin (1 : 1000), the 2nd leaf for the same period simply into Knop's solution. During all that time the leaves remained in darkness. The new determination gave the following values: $i_1=14.88$, $i_2=2.14$.

The experiments conducted by us enable one to the following conclusions.

Heteroauxin when introduced into the tissues of a green leaf in the form of a very weak solution (1 : 1000) causes a temporary increase in photosynthesis which attains a rather significant value: 100—200%. When longer action of this solution is applied, a drop of the photosynthetic energy below the normal level sometimes follow the rise observed at the beginning. Leaves immersed for several hours into a solution of β-indolyl-acetic acid of higher concentration (10 : 1000) have shown in our experiments neither noticeable increase nor decrease of the intensity of assimilation.

The increase of the energy of photosynthesis under the influence of injuries of the leaf may apparently be attributed to accumulation in the injured leaves of active substances (phytohormones) which act upon the process of assimilation like heteroauxin does.

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PLANT PHYSIOLOGY

MINIMUM INTERVALS OF DARKNESS AND LIGHT TO INDUCE FLOWERING IN SHORT DAY PLANTS

By B. S. MOSHKOV

(Communicated by N. I. Vavilov, Member of the Academy, 19. I. 1939)

It was in the year 1931 that the work of Garner and Allard(1) was published which completed the discovery of photoperiodism by showing that short day species, flowering rapidly at photoperiods of equal, 12-hour, night and day length, will not flower at all when the same dosage of light and darkness is alternated at shorter intervals, viz. from 6 hours to 5—3 seconds. On the other hand, long day species, which form no flowers when given 12-hour days, were found by those authors to flower when darkness and light were changed in quick succession.

With these factors two fundamental principles have been established for the first time: 1) the photoperiods have been interpreted properly as the daily ratio between the intervals of light and darkness, not as a daylength or nightlength independent of one another; 2) it has proved impossible to obtain a characteristic photoperiodic reaction in plants by breaking the nightly periods of their life, while an intermission of the daily period of light by darkness has been found to produce no influence on the flowering time of both short day and long day plants.

Very likely that part of the process which proceeds in the light either can be accomplished within very short periods of the day or is not intercepted by temporary darkness when the length of the latter does not exceed a certain limit, or at long last, is, as it were, summed up, i. e. advances by steps. But then another part of the process, which is shared to the night in the photoperiodic reaction, requires rather long daily periods of darkness which must not be interrupted by light. This part of the process is slow in progress and incapable of creating a summative effect nor to advance by degrees when interrupted.

In order to verify the above statements I started special experiments, whose results are set forth here. First of all I had to choose a material suitable for the purpose. Having studied the photoperiodic response of many plants I chose *Perilla ocymoides*. Most of the varieties of this species flower on daylengths of 3 to 16 hours. And there is no considerable difference between the photoperiods within this range as regards both budding and flowering.

On daylengths of 2 as well as 17 hours perilla fails practically to flower. The question therefore arises as to what is the minimum of the daily dosage of darkness and light indispensable for this species to pass the full cycle of development. It will be apparent that a daily period of light as long as 3—

4 hours is sufficient, since the plants fruit sparcely being given 3-hour day-lengths and satisfactorily when the days are lengthened to 4 hours. On the other hand, the minimum length of the dark period should not be reduced below 8 hours a day, because the plants will not flower when the length of the night does not exceed 7 hours, as it is the case with the 17-hour days. In the direction of long days the flowering of perilla is limited by 16-hour day alternated with 8-hour night. Its initiation under these conditions is somewhat delayed and fruiting is sparce even compared to 15-hour day when this species flowers profusely and quickly. Consequently, no less than 9 continuous hours of daily darkness are indispensable in order to make the seedlings of perilla fruit well. As our aim was to prove that the length of day and night requirement of perilla is as high as 3—4 hours of light and 8—9 hours of darkness while the rest of the daily period matters little, we subjected 20 day old plants of perilla to photoperiodic test for 20 days according to the plan referred to in the table. Before the start of the experiment and after the expiration of photoperiodic action all the plants tested enjoyed continuous illumination. To make use of the most long days the entire work was carried on in the hothouse during mid-summer. The control plants were kept under the conditions of continuous illumination throughout the experiment.

Furthermore there was a second set of control plants and these were given 12-hour days for the time when the experimental plants experienced their respective photoperiods (the 12-hour day may be accounted optimal for the development of perilla).

No of variant	Series	
	first	second
I	Photoperiods	
I	3 hours of light and 21 hours of darkness every 24 hours	4 hours of light and 20 hours of darkness every 24 hours
II	15 hours of light and 9 of darkness every 24 hours	16 hours of light and 8 hours of darkness every 24 hours
III	3 hours of light and 9 hours of darkness every 12 hours (i. e. two photoperiodic cycles a day)	4 hours of light and 8 hours of darkness every 12 hours (i. e. two photoperiodic cycles a day)

In all the experimental variants the first buds appeared almost simultaneously (with a difference of 2—3 days), 20—23 days after the plants had been subjected to photoperiods. Yet the further development of the buds varied according to the variant. In the first series, for instance (Fig. 1), the plants given 3-hour days produced vey few buds only on the top and these were yet closed at the moment the photograph was being taken. Obviously the plants of that variant were wanting light when preparing to flower. The plants of Variant II, given 15-hour days and 9-hour nights, were only slightly behind of the development of 12-hour controls and they fruited profusely (Fig. 1). Plants from Variant III were intermediate as regards development but approached closer to Variant I, i. e. suffered from light deficiency. From the behaviour of plants in the first series it would appear that 3 hours of light in the photoperiodic cycle present a minimum which cannot possibly be overstepped. The somewhat better development of the plants from Variant III compared to Variant I must, however, be taken to mean that either there exists some definite ratio between plant's needs in

darkness and light, or the processes taking place in the light increase by addition and end with producing a summative effect.

In the second series (Fig. 2) only plants from Variant I, which received 4-hour days, produced ripe fruits. Plants from 16-hour day variant were nearer to flowering than those given 3-hour days, but they did not start flowering either and formed few buds.

Fig. 1. *Perilla ocymoides*. First series. Variants I (left), II (right) and III (in the middle).

Plants of Variant III looked like those of Variant II. It will be obvious that 8 hours of darkness, whether following 16 or 4 hours of light, are insufficient for full development and profuse fruiting. It is only 9 hours of darkness and 4 hours of light which satisfy the requirement of perilla in photoperiodic conditions, and whether the remainder of 11 hours is shared to the day or to the night is of no consequence for the processes which prepare the

Fig. 2. *Perilla ocymoides*. Second series. Variants I (left), II (right) and III (in the middle).

plant to enter into reproductive stage. Yet the distribution of the rest of the daily period between light and darkness is not without influence on the yield of seed. Speaking of the minimum intervals of light and darkness in photoperiodic cycles suitable for the development of short day species, we can say that the length of the dark periods should be two and half or three times that of the light periods. There have been carried out

other experiments, which are not reported here, and they have shown tha minimum daylenghts can be broken by short intervals of darkness without detriment to the development of plants, while an intermission of minimum periods of darkness by light disturbs the development of short day plants, even if it does not change the totals of dark hours.

On the basis of the minimum requirement of plants in daily rations of darkness and light we were successful, in the experiments here described, for the first time to induce flowering in short day species by alternating light with darkness artificially in order to create two photoperiodic cycles instead of one within 24 hours.

Not only a theoretical significance is attached to the experiments reported; the latter have also pursued a practical purpose. This was namely to clear up the question whether the development of plants can be accelerated by narrowing the cycles of recurring external conditions to a limiting value. As early as in 1934 and 1935 the author of the present communication and somewhat later Potapenko at Michurinsk, Moroz at Sukhumi and Zelensky at Maikop attempted to accelerate the development of fruit-trees and some perennial woody species cultivated for industrial purposes by creating several years within one calendar year, these arfificial years including the chief seasons determining the development of plants: vegetation, transition to dormancy, lifting of dormant state, renewed vegetation. All those experiments furnished many interesting data to interprete the rôle of reduced temperature in lifting the dormant state as well as in subsequent growth, and the factors controlling the completion of growth and the fall of leaves came to light. The not indispensable nature of the dormant state was thereby established even for deciduous trees, but the experiments which aimed at accelerating the development of perennials failed to reach this end. In author's opinion, these results should be taken to mean that development is determined by definite external conditions of life and by time rather than by annual cycles. A quite another case is that of daily photoperiodic cycles in the life of short day plants; no doubt, their rôle in the development of perilla and its related species is very important. We failed, however, in our attempts to accelerate development by doubling those cycles within 24 hours, with the aim to create two days within the time of one calendar day. The fact that all our variants started budding simultaneously points toward nothing but that the daylengths ranging between 4 and 15 hours are the most favourable for the progress of certain part of developmental processes, though the latter defy acceleration by means of doubling the photoperiodic cycles within the same space of time. Every physiological process, aside from its requirements to environmental conditions, needs a definite minimum interval of time and will not come to completion until after this interval has elapsed. These are indeed rather complex phenomena we have to inquire into, but not without having solved them can we expect to obtain a real possibility of controlling the development of perennials.

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PLANT PHYSIOLOGY

PHOTOPERIODIC RESPONSE OF PLANTS AS DETERMINED BY THEIR ONTOGENESIS

By B. S. MOSHKOV

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Most of the works devoted to photoperiodism, not excepting those of Garner and Allard, have failed to examine the relationship between ontogeny and photoperiodic response of plants. Among early investigators it was Emerson([1]) who made closest approach to the problem, but he did not study it either. Later on Eghis ([2]), followed by Razumov, Lubimenko and Stscheglova, has shown that young plants respond to photoperiods as soon as 7—20 days after the photoperiodic action has set in. Their works gave rise to two new terms which were introduced into the literature on photoperiodism: photoperiodic after-effect (suggested by Maximov) and photoperiodic induction (suggested by Lubimenko and Stscheglova). Though approaching close to the question under review, these works, however, have neither set nor solved it. Nor has this been done by Dolgushin ([3]) who was working with millet, a material unsuitable for the purpose which has been a factor in his failure. Finally, Čajlachjan ([4]) set himself the task to find out the relation between the age of plants and their response to photoperiods, but the conclusion he came to is entirely wrong, as it will be shown further. A proper solution of the problem is, however, indispensable in order to further the work on acceleration of perennial plants development. With regard to citrus plants, for instance, as lemon, orange and mandarine, it has been established that whatever photoperiods their seedlings may be given, they will not flower under the conditions prevailing at Sukhumi and Pushkin until after they have become 2—3 years old. At the same time the plants of those species, obtained by the usual method of grafting, flower already within the second year after grafting when given short days and fail to flower even in the 4th year on continuous illumination. The facts referred to should be taken to mean that only after some preliminary age variations have taken place can the optimal photoperiodic conditions induce flowering. Similar results were reached in my experiments on chrysanthemum as early as 1933. After this plant is grown on continuous illumination and has then attained an age of 90 days, it requires no more than 5 short days to form the apical bud, while even 15 short days will not suffice, if the plant is only 45 days old (then it needs 20 short days). The witness borne by this fact is that even in vegetatively propagating plants age condition is of great importance, if flowering is their intended response to photoperiods. If cuttings taken at different times from a chrysanthemum clone are simultaneously put under short

day conditions, the following picture may be observed: 120, 90 and 60 day old cuttings start budding on 20—22nd day, whille 30 day old cuttings form buds only on 50—55th day. These facts alone would furnish sufficient evidence that an important factor in the response of plants to photoperiods is the age condition, in which they have to suffer the latter (ontogeny). Besides, as early as in 1936, the author has shown that the leaves on the same shoot of chrysanthemum vary in their response to photoperiods on account of their different physiological condition bound up with ontogeny.

At long last we can recur to the data obtained with an annual plant, perilla, on its propagation by seed. In the case of this early ripening plant the initial ontogenic condition proved also of great importance for the effect of optimal photoperiods. The results of the experiment on *Perilla ocymoides* are presented in Table 1. The seedlings varied in age between 45 day old plants and plants which had just appeared above ground, when they started to receive 10-hour days. Previously they were grown on continuous illumination. The seedlings which had not attained an age of 5 days when subjected to photoperiods began budding 20 days later. When 15 and more days old, they produced buds in 12 days.

Table 1

Date of curtailing light	Age of seedlings at the respective dates (germination, June 27)	Budding	
		Date	Number of days passed from the start of photoperiodic action
27.VI	1 day	19.VII	21
2.VII	5 »	22.VII	20
7.VII	10 »	24.VII	17
12.VII	15 »	24.VII	12
17.VII	20 »	29.VII	12
22.VII	25 »	3.VIII	12
27.VII	30 »	8.VIII	12
1.VIII	35 »	13.VIII	12
6.VIII	40 »	18.VIII	12
11.VIII	45 «	23.VIII	12

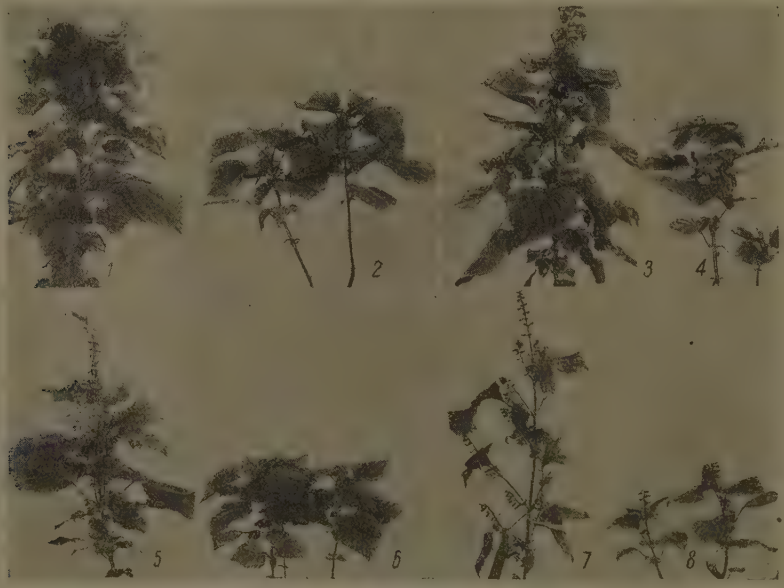
The data in Table 2 are the results of the second experiment on the same strain of perilla. In this experiment perilla enjoyed continuous illumination throughout its life, except for being given 10-hour light photoperiods during 7 days at different age, beginning with the stage of germination and up to the age of 45 days (through 5-day intervals). The seedlings ranging in age between one and 10 days failed to flower in response to the 7-day period of curtailed light, while those which attained an age of 15 days and more responded to the period of short days by flowering under the conditions of continuous illumination, the buds appearing morphologically in 18—22 days. The conclusion to be drawn from this experiment is that 7 days of optimal photoperiods will not suffice for 10 day old and younger perilla, while this time is quite sufficient for older plants, which have attained an age of 15 days.

These and some other experiments on *Perilla ocymoides* have shown that this plant under the conditions of Leningrad summer is ready to respond actively to the influence of optimal photoperiods, as soon as 7—10 days after seedlings have appeared above ground (germination). 7—10 days of optimal photoperiods suffice thereby to make the plants flower, and after 18—20 such days they fruit profusely.

Table 2

7 day period of curtailed light				Budding	
Time		Respective age (germination June 27)		Date	Number of days passed from the beginning of curtailed day period
from	to	from	to		
27.VI	4.VII	1	7	Nil	—
2.VII	9.VII	5	12	»	—
7.VII	14.VII	10	17	»	—
12.VII	19.VII	15	22	4.VIII	22
17.VII	24.VII	20	27	9.VIII	21
22.VII	29.VII	25	32	10.VIII	18
27.VII	3.VIII	30	37	16.VIII	19
1.VIII	8.VIII	35	42	19.VIII	18
6.VIII	13.VIII	40	47	24.VIII	18
11.VIII	18.VIII	45	52	29.VIII	18

On experimenting with another species of perilla (*Perilla nankinensis*) in 1937—38, we obtained still more clear-cut evidence that age condition of plants is a potent factor in their ability to avail themselves of the optimal photoperiods.



Perilla nankinensis. 50 days was the age of the odd numbered plants at the start of the short day period; 15 day, that of the even numbered plants. The first pair was given 10 short days; the second, 15 short days; the third, 20 short days and the fourth, 25 short days.

50, 15 and 5 day old seedlings of *Perilla nankinensis* being simultaneously given 5, 10, 15, 20, 25 and 30 curtailed days, it was established that: 1) 20 curtailed days are necessary to make this species fruit, 2) curtailing should be started 15—20 days after the plants appeared above ground, since if

begun earlier it fails to accelerate fruiting and moreover affects the green mass and hence the yield of seed unfavourably. This conclusion arose from the fact that 50 day old plants were usually observed to fruit only after they had received 20 short days (failing to flower after 15 such days). 15 day old plants being given 20 short days lagged far behind 50 day old ones (figure), but they came up with the latter after 25 short days (consequently, they had first to reach an age of 20 days). Finally, the 5 day old plants produced fruits as late as after 40 short days; it is therefore very likely that during the first 20 days they failed to accelerate their development in response to optimal photoperiods.

Perilla differs from millet in that its growing on continuous illumination gives no rise to processes which would convey it to flowering on short days, as it is the case with millet and most of the varieties of soya which are being prepared to flower on short days while growing on continuous illumination. The latter species are therefore unfit for the experiments which aim to reveal the rôle of ontogeny in the photoperiodic response of plants. This seems to be a factor in Čajlachjan's erroneous conclusion that «the discovery of photoperiodism has solved one of the principal problems in the domain of plant's physiological development by proving that the sexual maturity of plants is not definitely related to their age». It is just the contrary to what has been established by photoperiodic investigation, which has marked out quite distinct age stages indispensable for full sexual development of plants. It is a simple matter to infer from our examples of perennials and an annual plant (citrus plants and perilla) that ontogeny, which seems to be definitely related to the age condition of plants, essentially determines the rate of their development in response to optimal photoperiod. And not until the plants pass a certain stage of age or ontogeny, can they avail themselves of favourable photoperiods, but with this stage achieved, their age condition is no longer of any importance for their further development and fruiting.

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